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მოამბე

154

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1996

თბილისი * TBILISI

WE OFFER

A New Method of Production of Bulk SYNTHETIC DIAMONDS

of Electron and Positive Conduction

We offer to solve the problem of production of pure bulk synthetic diamonds of electron and positive conduction by means of phenolformaldehyde resin (FFR). The technology makes it possible to synthesize FFR purely and to form solid solution with Li, Br, P and any other chemical element. The concentration of introduced element is changed in a wide range $10^{15} \div 10^{19}$ cm^{-3}. Received solid solution is easily graphitized at the temperature of 2600-2800°C. The graphite with introduced element (Li, Br, P and others) of the necessary concentration is used as the basis for production of semiconductor diamond of N- and P- type. Diamond can be obtained by means of straight phase transformation of graphite. It can also be obtained by means of using graphite at high temperatures and static pressures in the presence of metal-solvents. By means of this method we can also get semiconductor graphite of P- and N- type.

Karumidze G.S.
I.Gverdtsiteli Institute
of Stable
Isotopes

Dundua V.U.
Institute of Physics of the
Georgian Academy
of Sciences

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52, Rustaveli Avenue,
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Republic of Georgia

Telephone : **+ 995 32 99.75.93**
Fax : **+ 995 32 99.88.23**
E-mail : **BULLETIN@PRESID.ACNET.GE**

forced, as he was asking the God David's death, knowing that David would have never given him the power [10]. And really the historian of Lasha-George time writes that after having taken monastic vows Demetre died in a year. but as we have seen before, Tamar's historian says that Demetre chose his son George for joint ruling. Iv. Dzavakhishvili concludes that "after taking monastic vows Demetre turned again to the royal life and set his favorite son George III on the throne for joint rulling [10].

The question of the XII century Georgian throne hereditary is not quite clear. David V evidently fights for the throne and by different measures taken it. The team of David's supporters is not known, but one is clear that it must have been rather strong, because in spite of Demetre's sympathy to his younger brother George, David could take the throne. His main supporter was Tirkash' son of Ivane Abuletisdze, who was killed by Demetre in 1132. Tirkash was the king's prisoner. After his accession to the throne, David set him free and appointed him to the Amirspasalar's post. By that he turned all the Amirspasalars of Orbel's family against himself. From Vardan's words because of that the Orbels (Sumbat and Ivane) connected with George III who had promised them to return the Amirspasalar posts and killed David the King [2]. Important notices about David's supporter feudal lords gives us Mkhitar Goshi in his 'Kroniks of Albania". The historian says that David has been on good terms with Vasak Iskhan-the satrap of Tbilisi and his brothers. David was so kindhearted that sent a man and invited Kvirike, King David Bagratuni's son and promised him to return his heredity, taken away from his ancestors. Having presented him in such a way he sent him back and agreed with him about their future meeting [15, p.46-47]. There is almost nothing known about Kvirike III. But this person especially attracts our attention in his connection with the royal court of Georgia in the second part of the XII century. Then historian countries: "when the Georgian nobles learned his such desire, the great envy rose in them, especially in those, who were named as Orbelians; they gave him poison and killed David the King. Hard and long sadness seized the Georgians and the Armenian countries" [15].

As we see David V looked for the numerical strength of his supporters not only in Georgians. With their help he tried to weaken his rival among Georgian nobles. It was known for David the great desire of the Armenians to restore Armenian State system. This desire had always been alive in their minds. In case of realization of this fact by the Georgian king's concept, the idea to restore Armenian State system would become more real. It is clear that David's interest in this question was not caused by his kindheartedness to the Armenians. His interest was from his own interests. It is no more chance that David choose Kvirike III Bagratun-Kvirikiani - the Lord of Matsnaberdi, he was direct heir of royal governors dynasty of Tashir-Dzoraget or Lore. David wanted to restore the kingdom of Lore for the purpose to weaken the Orbels. As to Orbels they took the place of Kvirikian Kings of Tashir-Dzoraget. David V's politics connected with the restorance of Armenian State system especially compromised the Orbels' interests. For a certain time they had lost the post of Amirspasalar and then there was the danger to loose the lands of Lore, because the restorance of former Tashir-Dzoraget kingdom of Kvirikians meant to leave the Orbels without lands. If all this is was happened the opposition between the Kvirikians and the Orbels would have been inevitable. There was no way for the Orbels but killing David (by Mkhitar Goshi). Varden's notice about David's killing by Orbels is confirmed by Mkhitar Goshi.

To avoid the conflict and to firm own position Orbels became related with Kvirikians. Ivane Orbeli married Kvirike III's daughter Rusugan. We suppose that this marriage would have happened after this silent conflict and David's death about in

CONTENTS

MATHEMATICS

CYBERNETICS

PHYSICS

ASTRONOMY

ORGANIC CHEMISTRY

HUMAN AND ANIMAL PHYSIOLOGY

BIOPHYSICS

BIOCHEMISTRY

ZOOLOGY

EXPERIMENTAL MEDICINE

EXPERIMENTAL MORPHOLOGY

LINGUISTICS

HISTORY

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

M.Bakuridze

Some Properties of Cesaro Means of Conjugate Trigonometric Series

Presented by Academician L.Zhizhiashvili, May 14, 1996

ABSTRACT. **Some results concerning the approximation properties of Cesaro means of conjugate trigonometric Fourier series of continuous functions are stated.**

1. We consider the 2π-periodic function of a real variable and denote $T=[-\pi,\pi]$. If the function $f\in L(T)$ then, as a rule, by $\sigma[f]$ and $\overline{\sigma}[f]$ we denote correspondingly Fourier trigonometric series of f and its conjugate series. By $t_n^\alpha(x,f)$ we denote Cesaro means of $\overline{\sigma}[f]$ of the order $\alpha>-1$, i.e.

$$t_n^\alpha(x,f)=\frac{1}{A_n^\alpha}\sum_{k=0}^{n}A_{n-k}^{\alpha-1}\overline{S_k}(x,f),$$

where $\overline{S_k}(x,f)$ is the k-th partial sum of the series $\overline{\sigma}[f]$ and

$$A_0^\alpha=1,\quad A_i^\alpha=\frac{(1+\alpha)(2+\alpha)\dots(i+\alpha)}{i!},\quad i\geq 1\quad (\alpha>-1).$$

Let $p\in[1,+\infty]$ be a certain number. If $f\in L^p(T)$ ($L^\infty(T)\equiv C(T)$) then by $\omega(\delta,f)_p$, $0<\delta\leq\pi$, we denote the L^p - modulus of f continuity.

2. In [1] we announced some properties of Cesaro means of the series $\sigma[f]$. They sharpen the corresponding results of Hardy and Littlewood [2], Yano [3] and others. In this paper some properties of the means $t_n^\alpha(x,f)$ are stated.

3. In the sequel it is supposed, that

$$\overline{f_n}(x)=-\frac{1}{2\pi}\int_{\pi/n}^{\pi}\left[f(x+t)-f(x-t)\right]ctg\frac{t}{2}dt,\quad n\geq 4.$$

The following is true:

Theorem 1. *Let $f\in C(T)$, and $\alpha\in]0,1[$ and $p\in]\frac{1}{\alpha},+\infty[$ be certain numbers. Then*

$$\left\| t_n^{-\alpha}(\cdot,f)-\overline{f_n}(\cdot)\right\|_{C(T)}\leq A(p,\alpha)\left[\omega\left(n^{\alpha-1},f\right)_C+n^\alpha\omega\left(\frac{1}{n},f\right)_p\right],$$

where $A(p,\alpha)$ is a positive finite constant, depending only on p and α parameters.

Theorem 2. *Let $f\in C(T)$, and $\alpha\in]0,1[$ and $\varepsilon\in]0,\frac{1}{2}[$ be certain numbers. If $\alpha p=1$, $\beta\in]0,\alpha[$, then*

$$\left\| t_n^{-\alpha}(\cdot,f)-\overline{f_n}(\cdot)\right\|_{C(T)}\le$$

$$\le A_1(p,\alpha)\left[\omega\left(n^{\alpha-1},f\right)_C+\omega^{1-2\varepsilon}\left(\frac{1}{n},f\right)_C+n^{\alpha}\omega\left(\frac{1}{n},f\right)_{1/\alpha}\omega^{\varepsilon(\alpha-\beta)}\left(\frac{1}{n},f\right)_C\right],$$

where $A_1(p,\alpha)$ is a positive finite constant, depending only on p and α parameters.

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

MATHEMATICS

I.Khasaia

Some Questions of Convergency of Spectral Expansions

Presented by Academician L.Zhizhiashvili, April 11, 1996

ABSTRACT. **For differential polynomial bundle with irregular disintegration boundary conditions, necessary conditions of uniform convergence of n–fold spectral expansions are accepted.**

Let's take differential bundle formed by differential equation

$$y^{(n)} + \sum_{j=1}^{n}\sum_{i=0}^{j} P_{ij}\lambda^{n-j} y^{(i)} - \lambda^n y = 0 \quad (P_{n,n} = 0) \tag{1}$$

and by normalized disintegration of boundary conditions

$$\begin{aligned} U_i(y) = U_{i0}(y) &= \sum_{l=0}^{k_i} \alpha_{i,l} y^{(l)}(0) = 0, \quad i = \overline{1,m}, \\ U_i(y) = U_{i1}(y) &= \sum_{l=0}^{\sigma_i} \beta_{i,l} y^{(l)}(1) = 0, \quad i = \overline{m+1,n}, \end{aligned} \tag{2}$$

$$n-1 \geq k_1 > k_2 > \ldots > k_m \geq 0,\ n-1 \geq \sigma_{m+1} > \sigma_{m+2} > \ldots > \sigma_n \geq 0,$$

$$Pij \in C^i\,[0,1]\,,\ 0 \leq i \leq j,\ 1 \leq j \leq n,\ Pjj = \text{const}.$$

Let $\omega_1, \omega_2, \ldots, \omega_n$ denote roots of characteristic equation
$\omega^n + P_{n-1,n-1}\,\omega^{n-1} + P_{n-2,n-2}\omega^{n-2} + \cdots + P_{2,2}\omega^2 + P_{1,1}\omega - 1 = 0$
Let $\omega_j = R_j e^{i\alpha_j}, R_j \neq 0, j = \overline{1,n},\ \alpha_j \neq \alpha_k,\ j \neq k$

λ plane is divided into lines $Re\lambda\omega_i = Re\lambda\omega_j$, $i \neq j$ in $2n$ sectors

$$S_\nu = \left\{\lambda, \gamma_\nu \leq \arg\lambda \leq \gamma_{\nu+1}\right\},\ \nu = \overline{0,2n-1},$$

where $0 \leq \gamma_0 < \gamma_1 < \ldots < \gamma_{2n-1} = \gamma_0 + 2\pi$.

For each sector S_ν, $\nu \in \{0,1,..,2n-1\}$ exist permutation "$p\nu$" of numbers $\{1,2,\ldots,n\}$ that for $\lambda \in S_\nu$ obtain

$$Re\lambda\omega_{p\nu(1)} \leq Re\lambda\omega_{p\nu(2)} \leq \ldots \leq Re\lambda\omega_{p\nu(n)} \tag{3}$$

Set of $\nu \in \{0,1,2,\ldots,2n-1\}$ are indicated with $\{\nu_1, \nu_2, \ldots, \nu_n\}$, where $\nu_i \neq \nu_j$, $i \neq j$.

Let's demand for each $\nu \in \{\nu_1, \nu_2, \ldots, \nu_n\}$

$$Re(e^{i\gamma\nu}\omega_{p\nu(m)}) > 0 \tag{4}$$

For each sector S_ν, $\nu = 0,1,\dots,2n-1$ exists the fundamental system of equation solution (1) ([1], p.66) $y_j(x, \lambda)$, $j = \overline{1,n}$ that for $\lambda \to \infty$ admits assymptotics

$$\frac{\partial^s}{\partial x^s} y_j = (\lambda\omega_j)^s e^{\lambda\omega_j x}\left\{\chi_j(x) + O\left(\frac{1}{\lambda}\right)\right\} \tag{5}$$

$x \in [0,1]$, $\lambda \in S_\nu$ function $\chi_j(x)$ may be shown in the aspect $\chi_j(x) = c_j \cdot e^{z_j \cdot x}$, where c_j, z_j – numbers differ from the nought. Suppose, that $\chi_j(0) = 1$, $1 \le j \le n$.

Let's indicate

$$\Omega^{(k)}_{p\nu(1),p\nu(2),\dots,p\nu(m)} = \begin{vmatrix} \omega^{k_1}_{p\nu(1)} & \cdots & \omega^{k_1}_{p\nu(m)} \\ \cdot & \cdot & \cdot \\ \omega^{k_m}_{p\nu(1)} & \cdots & \omega^{k_m}_{p\nu(m)} \end{vmatrix} \neq 0,$$

$$\Omega^{(\sigma)}_{p\nu(m+1),p\nu(m+2),\dots,p\nu(n)} = \begin{vmatrix} \omega^{\sigma_{m+1}}_{p\nu(m+1)} & \cdots & \omega^{\sigma_{m+1}}_{p\nu(n)} \\ \cdot & \cdot & \cdot \\ \omega^{\sigma_n}_{p\nu(m+1)} & \cdots & \omega^{\sigma_n}_{p\nu(n)} \end{vmatrix} \neq 0$$

Let's suppose

$$\left(\prod_{i=1}^{m}\alpha_{i,k_i}\right)\left(\prod_{i=m+1}^{n}\beta_{i,\sigma_i}\right) = a \neq 0,\ \sum_{i=1}^{m}k_i + \sum_{i=m+1}^{n}\sigma_i = \beta$$

From condition (2), using (5) the theorem of Laplas', we get for $\lambda \in S_\nu$ ($\nu = 0,1,\dots,2n-1$):

$$\Delta(\lambda) = \lambda^\beta a \prod_{i=m+2}^{n}\chi_{p\nu_j(i)}(1) e^{\lambda\sum_{i=m+1}^{n}\omega_{p\nu_j(i)}}\left\{\Omega^{(k)}_{p\nu_j(1),\dots,p\nu_j(m)}\Omega^{(\sigma)}_{p\nu_j(m+1),\dots,p\nu_j(n)} \times\right.$$
$$\times\left[\chi_{p\nu_j(m+1)}(1)\right] - \Omega^{(k)}_{p\nu_j(1),\dots,p\nu_j(m-1),p\nu_j(m+1)}\Omega^{(\sigma)}_{p\nu_j(m),p\nu_j(m+2),\dots,p\nu_j(n)} \times \tag{6}$$
$$\left.\times e^{\lambda(\omega_{p\nu_j(m)}-\omega_{p\nu_j(m+1)})}\left[\chi_{p\nu_j(m)}(1)\right] + O\left(\frac{1}{\lambda}\right)\right\}$$

And now it isn't difficult to prove the following theorem using [2].

Theorem 1. *In case of boundary problem (1)–(2) exist n infinite sequence of eigen numbers $\lambda_{k,j}$, $k \in N$ that in case of $|\lambda|$ there are enough adults satisfied with the following assymptotics*

$$\lambda_{k_1 j} = \frac{1}{\omega_{p\nu_j(m)} - \omega_{p\nu_j(m+1)}} \left\{ 2k\pi i + \pi i + \ln\theta_{\nu_i} + O\left(\frac{1}{k}\right) \right\}, \quad j = \overline{1,n} ,$$

where

$$\theta_{\nu_j} = -\frac{\Omega^{(k)}_{p\nu_j(1),\ldots,p\nu_j(m)} \Omega^{(\sigma)}_{p\nu_j(m+1),\ldots,p\nu_j(n)}}{\Omega^{(k)}_{p\nu_j(1),\ldots,p\nu_j(m-1),p\nu_j(m+1)} \Omega^{(\sigma)}_{p\nu_j(m),p\nu_j(m+2),\ldots,p\nu_j(n)}} \times \frac{\chi_{p\nu_j(m+1)}(1)}{\chi_{p\nu_j(m)}(1)}.$$

Besides begining from some numbers all proper numbers are simple.

From (6) using theorem 1 it's easy to prove the following lemma.

Lemma 1. *Let's indicate the domain with S^0_ν $\left(\nu = \overline{0,2n-1}\right)$ from S_ν by excluding the eigen numbers together with circular neighbourings, with one and the same rather small radius r. Then in case of $|\lambda|$*

$$\Delta(\lambda) \geq C|\lambda^\beta e^{\lambda \sum_{i=m+1}^{n} \omega_{p\nu_j(i)}}|, \quad \lambda \in S^0_\nu, \quad \nu = \overline{0,2n-1},$$

where C is constantly independent of λ.

Let's consider the function

$$\varphi(x,\lambda) = \begin{vmatrix} U_1(y_1) & U_1(y_2) & \cdots & U_1(y_n) \\ \cdot & \cdot & \cdot & \cdot \\ U_{n-1}(y_1) & U_{n-1}(y_2) & \cdots & U_{n-1}(y_n) \\ y_1(x,\lambda) & y_2(x,\lambda) & \cdots & y_n(x,\lambda) \end{vmatrix} \qquad (7)$$

From here, using (2), (5), the theorem of Laplas, and taking into account that for any $\lambda \in S_\nu$ $\left(\nu = \overline{0,2n-1}\right)$ inequality (3) takes place, we'll get

$$\varphi(x,\lambda) = \left(\prod_{i=1}^{m} \alpha_{i,k_i}\right)\left(\prod_{i=m+1}^{n-1} \beta_{i,\sigma_i}\right) \lambda^{\sum_{i=1}^{m} k_i + \sum_{i=m+1}^{n-1}\sigma_i} \prod_{i=m+1}^{n-1} \chi_{p\nu_j(i)}(1) e^{\lambda \sum_{i=m+2}^{n} \omega_{p\nu_j(i)}} \times$$

$$\times \left\{ \sum_{l=1}^{m} e^{\lambda\omega_l x}\left[\chi_l(x)\right](-1)^{n+l} \Omega^{(k)}_{p\nu_j(1),\ldots,p\nu_j(l-1),p\nu_j(l+1),\ldots,p\nu_j(m+1)} \times \right.$$

$$\times \Omega^{(\sigma)}_{p\nu_j(m+2),\ldots,p\nu_j(n)} + e^{\lambda\omega_{m+1}x}\left[\chi_{m+1}(x)\right](-1)^{m+n+1} \Omega^{(k)}_{p\nu_j(1),\ldots,p\nu_j(m)} \times$$

$$\times \Omega^{(\sigma)}_{p\nu_j(m+2),\ldots,p\nu_j(n)} + \sum_{l=m+2}^{n} e^{\lambda\omega_l x}\left[\chi_l(x)\right] e^{\lambda(\omega_{p\nu_j(m+1)} - \omega_{p\nu_j(l)})} \times$$

$$\left. \times \Omega^{(k)}_{p\nu_j(1),\ldots,p\nu_j(m)} \Omega^{(\sigma)}_{p\nu_j(m+1),\ldots,p\nu_j(l-1),p\nu_j(l+1),\ldots,p\nu_j(n)} \right\} \qquad (8)$$

Let's indicate

$$a_1 = a/\beta_{n,\sigma_n}\ , \quad \beta_1 = \beta - \sigma_n .$$

Let's define the function $\psi(x,\lambda)$ in the following way

$$\psi(x,\lambda) = a_1^{-1}\lambda^{-\beta_1} e^{-\lambda \sum\limits_{i=m+2}^{n} \omega_{pv_j(i)}} \left(\prod_{i=m+1}^{n-1} \chi_{pv_j(i)}(1) \right)^{-1} \varphi(x,\lambda). \tag{9}$$

From here and (7) follows, that $\psi(x,\lambda_{k_1 j})$ eigen function confirms with eigen numbers $\lambda = \lambda_{k_1 j}$

From (9), taking (8) and (6) into account, we'll get

$$\psi(x,\lambda) = e^{\lambda\omega_{pv_j(m+1)}x}\left[\chi_{pv_j(m+1)}(x)\right](-1)^{m+n+1}\Omega^{(k)}_{pv_j(1),\dots,pv_j(m)}\,\Omega^{(\sigma)}_{pv_j(m+2),\dots,pv_j(n)} +$$

$$+(-1)^{m+n} e^{\lambda\omega_{pv_j(m)}x}\left[\chi_{pv_j(m)}(x)\right]\Omega^{(k)}_{pv_j(1),\dots,pv_j(m-1)}$$

From here, theorem 1, using [3], shows that it isn't difficult to prove the following lemmas

Lemma 2. *There exist constant $C > 0$ not depending on k and q,*

$$\left|\frac{d^q}{dx^q}\psi(x,\lambda_{k_1 j})\right| \le C\left|\lambda_{k_1 j}\right|^q \left|e^{\lambda_{k_1 j}\omega_{pv_j(m+1)}x}\right|, \quad 0 \le q \le n- \ .$$

Lemma 3. *Let [c,d] be arbitrary segment from interval (0,1). Then k_0 exists, depending on [c,d] that for every $k > k_0$ there will be $x_k \in [c,d]$ that*

$$\left|\psi(x,\lambda_{k_1 j})\right| \ge C\left|e^{\lambda_{k_1 j}\omega_{pv_j(m+1)}x_k}\right|, \quad 1 \le j \le n,$$

where C is constantly independent on k.

According to theorem 1 we've got n series eigen number $\lambda_{k_1 j}$ $\left(j = \overline{1,n}\right)$. Renumbering all the eigen numbers in increasing order of magnitude, let's indicate them λ_k (k =1,2,...).

So beginning from some numbers all eigen numbers are simple, thus without loss of generality we may regard, that to each λ_k (k =1,2,...) confirms the only eigen function (if λ_k is not simple, then the final numbers of joined elements are added). Let $\psi(x,\lambda_1)$, $\psi(x,\lambda_2)$,... be conformable system to eigen function.

Let's take into consideration (n−1) system of the elements $\lambda_1^{\nu}\psi(x,\lambda_1), \lambda_2^{\nu}\psi(x,\lambda_2),\dots$ $\nu = 1,2,\dots,n-1$.

This system of elements is called derivative chain according to Keldish [4].

Definition. If in case of some coefficients $\left\{a_k\right\}_{k=1}^{\infty}$, independent on ν, the following equality is accepted

$$\sum_{k=1}^{\infty} a_k \lambda_k^{\nu} \psi(x,\lambda_k) = f_{\nu}, \quad 0 \le \nu \le n-1$$

for given n function $f_0, f_1, \ldots, f_{n-1}$ we'll say n–fold expansion takes place in terms of eigen and joined functions of boundary problem (1)–(2).

Theorem has its place.

Theorem 2. *If one of these series*

$$\sum_{k=1}^{\infty} a_k \lambda_k^{s} \psi(x,\lambda_k), \quad (s = np+\nu,\ p = 0,1,\ldots,\ \nu = \overline{0,n-1}) \tag{10}$$

uniformly converges on $[c,d] \subset (0,1)$, then

1) series (10) converge absolutely and uniformly on $[0,a]$ $(a<d)$;

2) if we indicate by f_ν $(\nu = 0,1,\ldots,n-1)$ sums of the first n series (10), then the sums of every following n series is determined by the formula

$$f_s = f_{s-n}^{(n)} + \sum_{j=1}^{n} \sum_{i=0}^{j} P_{ij}\, f_{s-j}^{(i)} \quad (s \ge n)$$

3) if $x \in [0,d-\varepsilon]$, $0 < \varepsilon < d$, $C > 0$, depending on ε then f_s functions admit valuations

$$\left| \frac{d^l}{dx^l} f_{np+\nu} \right| \le C \left(\frac{1+\varepsilon}{\left| R_{p\nu_j(m+1)} \right| \cos\alpha_{p\nu_j(m+1)} (d-\varepsilon-x)} \right)^{np+l+\nu} (np+l+\nu)!$$

$(l = 0,1,\ldots,n-1;\ p = 0,1,2,\ldots;\ \nu = 0,1,\ldots,n-1)$.

4) f_s $(s = np+\nu,\ p=0,1,\ldots)$ functions satisfy boundary conditions in zero.

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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MATHEMATICS

U.Goginava

About Uniform Convergence and Summability of Fourier–Walsh–Paley Series

Presented by Academician L.Zhizhiashvili, June 3, 1996.

ABSTRACT. **The theorems about the uniform convergence and summability by negative order Cesaro method for multiple Fourier–Walsh–Paley series convergence defining in Pringshein sense and final assertions in some classes of functions have been determined.**

Below we use the following definitions and notations of [6].

Let us assume $M = \{1,2,...,N\}$, $B = \{l_1,...l_r\} \subset M$, $l_k < l_{k+1}$, $k=\overline{1,r-1}$, $i_B = \left\{i_{l_1},...,i_{l_r}\right\}$ where $i_{l_k} \in M, k=\overline{1,r}$. Then let $m = (m_1,...,m_N)$, $\alpha = (\alpha_1,...,\alpha_N)$, where m_k –natural numbers, and $\alpha_k \in]-1,0[$, $k=\overline{1,N}$. For the whole number m we shall denote vector $(m,...,m)$ of space E_N by symbol $\widetilde{m}$. Addition of m, α, $\widetilde{m}$ vectors we shall understand in the sense of Euclidean space E_N operation.

We shall suppose that $M(s)$ – set of all M subsets $M'(s) = M(s) \setminus \{\varnothing\}$. Later on we shall identify symbols

$$\sum_{i_B=\widetilde{1}}^{m} \quad \text{and} \quad \sum_{i_{l_1}=1}^{m_{l_1}} \cdots \sum_{i_{l_r}=1}^{m_{l_r}} ;$$

then record $m_B \to \infty$ means that $m_{l_i} \to \infty$, $l_i \in B$. Let us assume f be continuous on $Q_N = [0,1]^N$, i periodical relative to each variable.

Let

$$\Omega_{l_1} = f(x_1,...,x_{l_1}^{(1)},...,x_N) - f(x_1,...,x_{l_1}^{(0)},...x_N).$$

Expression which is obtained by succesive application of operations $\Omega_{l_1},...,\Omega_{l_r}$ is denoted by symbol Ω_B.

Implement following function (when $N = 1$, see [7])

$$\varphi_B(m_B,\delta_B,f) = \sup_{x_{M\setminus B}} \sup_{\Pi_{m_B\delta_B}} \sum_{i_B=\overline{1}_B}^{m_B} \omega_B(f, I_{i_B}^{(B)}),$$

where

$$\omega_B(f, I_{i_B}^{(B)}) = \sup_{x_B^{(1)},x_B^{(0)} \in I_{i_B}^{(B)}} \left|\Omega_B\right|,$$

and $\Pi_{m_B\delta_B}$ – system from $m_B = m_{l_1} \cdots m_{l_r}$ noncrossing intervals

$$\left\{I_{i_B}^{(B)}\right\}_{i_B=\tilde{1}_B}^{m_B} = \left\{I_{i_{l_1}}^{(l_1)}\times\cdots\times I_{i_{l_r}}^{(l_r)}\right\}_{i_{l_1},\dots,i_{l_r}=1}^{m_{l_1},\dots,m_{l_r}}$$

of the segment $[0,1]^r$,

$$\left|I_{i_{l_k}}^{(l_k)}\right| = \delta_{l_k}\ ,\ i_{l_k} = \overline{1,m_{l_k}}\ ,\ k=1,2,\dots,r.$$

Let $\varphi(k,\ \delta)$ –arbitrary function of the number k and non–negative δ satisfying following conditions:

1) $\varphi(k,\ 0) = \varphi(0,\ \delta) = 0$, $k = 0,1,\dots$, $\delta \geq 0$;
2) $\varphi(k,\ \delta)$ continuous and not decreasing relative to δ;
3) $\varphi(k,\ \delta)$ up–convex and not decreasing relative k;
4) $\varphi(k,\ \delta) \leq c\ \varphi\left(\left[k\cdot\frac{\delta}{\eta}\right],\eta\right)$, $\delta \geq \eta$;

c – some constant.

$M(\varphi)$ denotes set of all those f functions for which condition

$$\varphi(k,\ \delta,f) = O\ (\varphi(k,\ \delta))$$

is fulfilled.

Let $S_m(f,\ x)$ be rectangular particular sum of N–multiple Fourier–Walsh–Paley series of function f,

$$S_m^{\alpha}(f,x) = \frac{1}{A_m^{\alpha}}\sum_{i_M=\tilde{O}_M}^{m} A_{m-i_M}^{\alpha-\tilde{1}}\ S_{i_M}(f,x)$$

- Cesaro $(c,\ \alpha)$ – means of particular sums $S_m(f,\ x)$.

We prove the following theorems.

Theorem 1.

1) If function $f \in C(Q_N)$ and

$$\sum_{B\in M'(s)}\sum_{i_B=\tilde{1}_B}^{m_B}\frac{\varphi_B(i_B,1/m_B,f)}{i_B^2} \to 0\quad (m_M \to \infty),$$

then

$$\lim_{m\to\infty}\left\| S_m(f)-f\right\|_{C(Q_N)} = 0,$$

2) If condition

$$\overline{\lim_{n\to\infty}}\sum_{i=1}^{n}\frac{\varphi(i,1/n)}{i^2} > 0$$

is fulfilled, then in $M(\varphi)$ class continuous function will exist Fourier–Walsh–Paley series $S_m(f_0,\ x)$ of which divergences at point $x = 0$.

Theorem 2.

1) If for function $f \in C(Q_N)$ *condition*

$$\sum_{B \in M'(s)} \sum_{i_B=\bar{1}_B}^{m_B} \frac{\varphi_B(i_B, 1/m_B, f)}{i_B^{\bar{2}-\alpha}} \to 0 \quad (m_M \to \infty)$$

is fulfilled then

$$\lim_{m\to\infty} \left\| S_m^{-\alpha}(f) - f \right\|_{C(Q_N)} = 0.$$

2) If condition

$$\overline{\lim_{n\to\infty}} \sum_{i=1}^{n} \frac{\varphi(i, 1/n)}{i^{2-\alpha}} > 0$$

is fulfilled, then in class $M(\varphi)$ *continuous function will exist for which Cesaro means* $S_m^{-\alpha}(f_0, 0)$ *will diverge at the point.*

Theorems 1 and 2 contain in particular settlements work [8–10] and are spread in wider classes of function [11].

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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MATHEMATICS

O.Jokhadze

The First Mixed Problem for Pseudoparabolic Equations on a Plane

Presented by Academician T.Burchuladze, July 23, 1996

ABSTRACT. **The first mixed problem for the third order equations with dominated lower terms is considered in case of two independent variables. Several sufficient conditions and also one necessary and sufficient condition for the unique solvability of the considered problem are obtained. The example is constructed, which shows that if one of these sufficient conditions are not valid, then uniqueness of the problem solution can in general fail.**

In the plane of independent variables x,y let us consider partial differential equation of the third order

$$u_{xxy} + a^{2,0}u_{xx} + a^{1,1}u_{xy} + a^{1,0}u_x + a^{0,1}u_y + a^{0,0}u = f, \tag{1}$$

where $a^{i,j}$, $i=0,1,2,$ $j=0,1,$ $i+j\neq 3,$ f are the given functions and u is the unknown real one.

Straight lines y = const form a double family of characteristics of (1), while x = const a single family. Let $D \equiv \{(x,y)\colon 0< x < x_0,\ 0 < y < y_0\}$. For the equation (1) we shall consider the first mixed boundary value problem formulated as follows: Find in $\overline{D}$ a regular solution u of the equation (1) , satisfying the following initial–boundary value conditions

$$\begin{gathered} u(x,0) = f_1(x),\ 0 \le x \le x_0, \\ u(0,y) = f_2(y),\ u(x_0,y) = f_3(y),\ 0 \le y \le y_0, \end{gathered} \tag{2}$$

where $f_1 \in C^2[0,x_0]$, $f_2, f_3 \in C^1[0,y_0]$– are the given real functions, and $f_1(0) = f_2(0)$, $f_1(x_0) = f_3(0)$.

A function u, continuous in $\overline{D}$ together with its partial derivatives $D_x^i D_y^j$ u,i =0,1,2, j = 0,1, i+j > 0, $D_x \equiv \frac{\partial}{\partial x}, D_y \equiv \frac{\partial}{\partial y}$ and satisfying the equation (1) in D is called a regular solution of the equation (1).

It should be noted that the initial–boundary value problems for a wide class of equations of the third and higher order with dominated derivatives are treated in [1]–[3] and other papers.

Considering (1), (2) in the space $C^{2,1}(\overline{D}) \equiv \{u\colon D_x^i D_y^j u \in C(\overline{D}), i=0,1,2, j=0,1\}$ it is required that $a^{i,j} \in C^{i,j}(\overline{D})$, $i=0,1,2,$ $j=0,1,$ $i+j\neq 3,$ $f \in C(\overline{D})$.

According to the papers [2, 3] Rieman function $\upsilon(x,y;\xi,\eta)$ for the equation (1) is defined as a solution of the Goursat problem with the following conditions

$$-\upsilon_{xyx}+(a^{2,0}\upsilon)_{xx}+(a^{1,1}\upsilon)_{xy}-(a^{1,0}\upsilon)_x-(a^{0,1}\upsilon)_y+a^{0,0}\upsilon=0,$$

$$\upsilon(\xi,y;\xi,\eta)=0,\ \upsilon_x(\xi,y;\xi,\eta)=\exp\left\{\int_\eta^y a^{2,0}(\xi,y_1)dy_1\right\},\ \upsilon(x,\eta;\xi,\eta)=\omega(x,\eta;\xi,\eta),$$

where $\omega(x,\eta;\xi,\eta)$ – is the solution of the ordinary linear differential equation of the second order

$$\upsilon_{xx}(x,\eta;\xi,\eta)-(a^{1,1}\upsilon)_x(x,\eta;\xi,\eta)+(a^{0,1}\upsilon)(x,\eta;\xi,\eta)=0 \qquad (3)$$

with the initial conditions of the Cauchy $\upsilon(\xi,\eta;\xi,\eta)=0$, $\upsilon_x(\xi,\eta;\xi,\eta)=1$.
It is well known [2, 3] that the above defined Rieman function $\upsilon(x,y;\xi,\eta)$ is uniquely determined, also $D_x^iD_y^j\upsilon, D_\xi^iD_\eta^j\upsilon\in C(\overline{D}\times\overline{D})$, $i=0,1,2$, $j=0,1$, and for the regular solution u of the equation (1) the following integral representation

$$u(x,y)=\int_0^x\Big[\upsilon_x(x_1,0;x,y)\varphi_1'(x_1)+(a^{1,1}\upsilon)_x(x_1,0;x,y)\varphi_1(x_1)-$$

$$-(a^{0,1}\upsilon)(x_1,0;x,y)\varphi_1(x_1)\Big]dx_1-\int_0^y\Big[\upsilon(0,y_1;x,y)\nu_1'(y_1)+(a^{2,0}\upsilon)(0,y_1;x,y)\nu_1(y_1)\Big]dy_1-$$

$$-\int_0^y\Big[\upsilon_{xy}(0,y_1;x,y)-(a^{2,0}\upsilon)_x(0,y_1;x,y)-(a^{1,1}\upsilon)_y(0,y_1;x,y)+ \qquad (4)$$

$$+(a^{1,0}\upsilon)(0,y_1;x,y)\Big]\psi_1(y_1)dy_1-(a^{1,1}\upsilon)(0,y;x,y)\psi_1(y)+\upsilon_x(0,y;x,y)\psi_1(y)+$$

$$+(a^{1,1}\upsilon)(0,0;x,y)\varphi_1(0)-\int_0^x\int_0^y\upsilon(x_1,y_1;x,y)f(x_1,y_1)dx_1dy_1$$

is valid, where $\varphi_1(x)\equiv u(x,0)$, $0\le x\le x_0$, $\psi_1(y)\equiv u(0,y)$, $\nu_1(y)\equiv u_x(0,y)$, $0\le y\le y_0$.

To investigate the problem (1), (2) we use integral representation (4) and substitute $x=x_0$. After partial integration in the right–hand second integral we get the problem (1), (2) equivalent to the third kind Volterra integral equation

$$\upsilon(0,y;x_0,y)\nu_1(y)-\int_0^y\Big[\upsilon_y(0,y_1;x_0,y)-(a^{2,0}\upsilon)(0,y_1;x_0,y)\Big]\nu_1(y_1)dy_1= \qquad (5)$$

$$=g_1(y),\quad 0\le y\le y_0,$$

where g_1 is the known function.

Due to $f_2, f_3 \in C^1[0, y_0]$ and the properties of the Rieman functions $\upsilon(x, y; \xi, \eta)$ we can conclude that the right–hand side and kernel of the integral equation (5) is continuously differentiable with respect to the variable $y \in [0, y_0]$. Thus it is sufficient to investigate the equation (5) in the class $C[0, y_0]$ as any solution on this equation from the class $C[0, y_0]$ automatically belongs to $C^1[0, y_0]$. To show the existence of a unique solution of the integral equation (5) in the class $C[0, y_0]$ it is sufficient to show that $\upsilon(0, y; x_0, y)$ differs from zero everywhere on the segment $0 \le y \le y_0$. The following is proved.

Theorem 1. *If one of these conditions is fulfilled*

$$a^{0,1}(x,y) \le a_x^{1,1}(x,y),\ a^{0,1}(x,y) \le \frac{1}{2} a_x^{1,1}(x,y),\ a^{0,1}(x,y) \le 0,\ (x,y) \in \overline{D}, \qquad (6)$$

then the problem (1), (2) has a unique solution in the class $C^{2,1}(\overline{D})$.

The following example shows that if for example the second condition from (6) is not valid, then generally uniqueness of the problem solution (1), (2) fails.

Indeed

$$u(x,y) = \int_0^y \exp\left\{-\frac{1}{2}\int_0^x a^{1,1}(\xi, y)\,d\xi\right\} \sin\frac{\pi}{x_0}x,\ (x,y) \in \overline{D},$$

is the solution of the following equation

$$u_{xxy} + a^{1,1}u_{xy} + a^{0,1}u_y = 0,\ (x,y) \in \overline{D},$$

with the homogeneous initial–boundary conditions (2). Here

$$a^{0,1}(x,y) = \frac{1}{2}a_x^{1,1}(x,y) + \frac{1}{4}\left[a^{1,1}(x,y)\right]^2 + \left(\frac{\pi}{x_0}\right)^2,\ (x,y) \in \overline{D}.$$

One sufficient condition is also suggested while the condition (6) may fail, but the problem (1), (2) is uniquely solvable in the space $C^{2,1}(\overline{D})$. The following is valid.

Theorem 2. *If the following condition*

$$a^{0,1}(x,y) = \frac{1}{2}a_x^{1,1}(x,y) + \frac{1}{4}\left[a^{1,1}(x,y)\right]^2,\ (x,y) \in \overline{D} \qquad (7)$$

is valid, then the problem (1), (2) is uniquely solvable in the space $C^{2,1}(\overline{D})$.

Remark 1. It is obvious that another sufficient conditions of the type (7), providing unique solvability of the problem (1), (2) can be obtained on the base of the effective representation of the solution of the equation (3).

Equivalent formulation of the unique solvability of the problem (1), (2) is given in terms of one spectral problem for an ordinary differential equation of the second order.

Denote by $\Lambda(y), 0 \le y \le y_0$ the set of all eigenvalue λ of the following Dirichlet problem of the operator l

$$l\omega = \lambda\omega,\ \omega(0,.) = \omega(x_0,.) = 0,$$

where $(l\omega)(x,.) \equiv \omega_{xx} - (a^{1,1}\omega)_x(x,.) + (a^{0,1}\omega)(x,.), 0 \le x \le x_0$.

The following is valid

Theorem 3. *The problem (1), (2) is uniquely solvable in the space* $C^{2,1}(\overline{D})$ *for any* $f \in C(\overline{D})$, $f_1 \in C^2[0,x_0]$, $f_i \in C^1[0,y_0]$, $i = 2,3$, *if and only if* $0 \notin \Lambda(y)$, *for all* $0 \le y \le y_0$.

Apply the Theorem 3 to the operator l, with constant coefficients relative to the variable x, which corresponds to the case when the coefficients $a^{1,1}$ and $a^{0,1}$ of the equation (1) do not depend on $x(a^{1,1}(x,.) \equiv a^{1,1}(.), a^{0,1}(x,.) \equiv a^{0,1}(.))$.

Denote $\Delta(y) \equiv [a^{1,1}(y)]^2 - 4a^{0,1}(y)$, $0 \le y \le y_0$, $N \equiv \{1,2,...,\}$. By the Theorem 3 we have the following

Theorem 4. *Let the coefficients* $a^{1,1}$ *and* $a^{0,1}$ *in the equation (1) do not depend of the variable x. Then the boundary value problem (1), (2) is uniquely solvable in the space* $C^{2,1}(\overline{D})$, *for all* $f \in C(\overline{D})$, $f_1 \in C^2[0,x_0]$, $f_i \in C^1[0,y_0]$, $i = 2,3$, **if and only if the set**

$$I \equiv \left\{ y \in [0, y_0]: \Delta(y) < 0, \frac{x_0\sqrt{-\Delta(y)}}{2\pi} \in N \right\}$$

is empty.

Finally, let us give one more sufficient condition for unique solvability of the problem (1), (2) in the space $C^{2,1}(\overline{D})$.

Let the following unequality

$$kx_0 \exp\{kx_0\} < 2 \tag{8}$$

be valid, where $k \equiv \max_{(x,y)\in\overline{D}} \left\{ \left|a^{1,1}(x,y)\right| + x\left|a_x^{1,1}(x,y) - a^{0,1}(x,y)\right| \right\}$.

The following is valid

Theorem 5. *If the condition (8) is fulfilled, then the problem (1), (2) is uniquely solvable in the space* $C^{2,1}(\overline{D})$, *for all* $f \in C(\overline{D})$, $f_1 \in C^2[0,x_0]$, $f_i \in C^1[0,y_0]$, $i = 2,3$.

Remark 2. It is obvious that one can obtain in a similar way some other sufficient conditions of the type (8).

Remark 3. Likewise one can consider the second mixed boundary value problem for the equation (1), when for $x = 0$ derivative value in the direction of the normal of the solution is given, i.e., $u_x(0,y) = f_4(y)$, $0 \le y \le y_0$.

A.Razmadze Mathematical Institute
Georgian Academy of Sciences

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MATHEMATICS

K.Bitsadze

On Divergence of Multiple Fourier–Haar Series of Bounded Function of Several Variables on Zero Measure Set

Presented by Academician L.Zhizhiashvili, June 18, 1996

ABSTARCT. **For arbitrary zero measure subset E of h–dimensional cube $[0,1]^n$ there exists a bounded measurable function f of $[0,1]^n$ so that for each point of E and arbitrary number $\lambda>1$ the n–fold Fourier–Haar series of f function diverges on the λ.**

The problems of divergence for trigonometric series have been the subject of investigation for a long time [1–11].

Analogous problems were investigated for Haar series. We consider some of them being connected with the problem presented in this paper. Historical information and problems concerning Haar series are given in the papers of B.I.Golubov [12], P.L.Uljanov [13], and W.R.Wade [14].

The Fourier–Haar series of a Lebesgue integrable function f (on $[0,1]$) has the form:

$$f(t) \sim \sum_{n=1}^{\infty} a_n(f) X_n(t), \quad t \in [0,1], \tag{1}$$

where $a_n(f) = \int_0^1 f(t) X_n(t)\,dt, \quad n = 0,1,2,\ldots,$ are the Fourier–Haar coefficients of the function f, $\{X_n\}_{n-1}^{\infty}$ is the Haar system of functions [15].

A.Haar [16] proved that for each function f, continuous on $[0,1]$, Fourier–Haar series (1) converging to f, is uniformly on $[0,1]$ [15]. He [16] also proved that for each Lebesgue integrable function f the series (1) converges to $f(t)$ for almost every $t \in [0,1]$ [15].

V.I.Prokhorenko [17] proved that for each set of zero measure there exists a function $f \in \bigcap_{p\geq 1} L^p$ whose Fourier–Haar series diverges.

M.A.Lunina [18] proved that of φ is any even function nondecreasing on $[0, \infty)$ with $\varphi(0)=0$, $\lim_{x\to+\infty} \varphi(\infty) = +\infty$, that for each set E of type $\widetilde{G}_\delta$ and of zero measure there exists a function $f \in L \bigcap \varphi(L)$ whose Fourier–Haar series diverges unboundedly on E and converges on $[0,1]\backslash E$ [18].

V.M.Bugadze [19] proved that for each set of zero measure there exists a bounded measurable function whose Fourier–Haar series diverges on that set.

We proved theorem (formulated below) which transfers the latter result to the multiple Fourier–Haar series of the functions of several variables (for the convergence in the sense of Pwingsheim; moreover for λ–convergence).

The n–fold Fourier–Haar series of a Lebesgue integrable function f, defined on the n–dimensional cube $[0,1]^n$, has the form:

$$f(t_1,\dots,t_n) \sim \sum_{m_1,\dots,m_n=1}^{\infty} a_{m_1,\dots,m_n}(f)\, X_{m_1}(t_1)\dots X_{m_n}(t_n), \qquad (2)$$

$$(t_1,t_2,\dots,t_n)\in[0,1]^n,$$

where $a_{m_1,\dots,m_n}(f) = \int\limits_{[0,1]^n} f(t_1,\dots,t_n) X_{m_1}(t_1)\cdots X_{m_n}(t_n)dt_1\cdots dt_n$

are the Fourier–Haar coefficients of the function f, and $\left\{X_{m_1}\cdots X_{m_n}\right\}$ is the n–fold Haar system, $m_1,\dots,m_n=1,2,\dots$.

Theorem. *For arbitrary subset E of zero measure of n–dimensional cube $[0,1]^n$ there exists a bounded measurable function f given on $[0,1]^n$ such that for each point of the set E and arbitrary number $\lambda>1$ the n–fold Fourier–Haar series of a function f diverges on the subset E on the λ–converges opinion.*

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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MATHEMATICS

T.Tevzadze

Multiple Fourier Series with Nonnegative Coefficients

Presented by Academician L.Zhizhiashvili, June 10, 1996

ABSTRACT. **The theorem with generalized results of M. and Sh. Izumi and of Boas for functions of several variables is stated.**

1. We denote by $x=(x_1,...,x_k)$, $y=(y_1,...,y_k)$, $t=(t_1,...,t_k)$,..., elements of k–dimensional ($k\geq1$) Euclidian space. We shall mean that $M=\{1,2,...,k\}$, $B\subset M$, $B'=C_mB$, $T^k=[-\pi,\pi]^k$ and $R^k(B)$ is hyperplane spanned over those and only those coordinates of vector x which forms a set B. We also mean that

$$H^k=]0,1]^k,\ H_1(t,B)=]0,t_i[^k\cap R^k(B),\ i\in B,$$
$$H_2(t,B)=[t_j,1]^k\cap R^k(B),\ j\in B.$$

A function under consideration f is 2π– periodic with respect to each coordinate of vector x and $f\in C(T^k)$.

As usual

$$\Delta_m f(x,t)\equiv\sum_{i=1}^{m}(-1)^{\sum_{j=1}^{k} i_j}\prod_{j=1}^{k}\binom{m_j}{i_j}f[x+(m-2i)t.$$

We consider such continuous function of k variables whose k–fold trigonometric Fourier series have the following form:

$$\sigma_k|f|\equiv\sum_{n\geq0}2^{-\lambda(n)}a_n(f)\prod_{i\in B}\cos n_i\cdot x_i\prod_{j\in B'}\sin n_jx_j,\qquad(1)$$

where $\lambda(n)$ is the number of all coordinates equal to zero. We mention, that if $B'=\varnothing$ in (1), then we have k–fold "cosine" series and if $B'=M$ (i.e. $B=\varnothing$) then we have k–fold "sine" series.

2. In case of one dimension behavior of series

$$\frac{a_0(f)}{2}+\sum_{n=1}^{\infty}a_n(f)\cos nx,\ \ x\in]-\infty,\infty[,\qquad(2)$$

and

$$\sum_{i=1}^{\infty}b_n(f)\sin nx,\ \ x\in]-\infty,\infty[\qquad(3)$$

is well investigated. This is confirmed by fundamental statements given in the monographs of Bari [1] and Zigmund [2] and Pak's paper [3].

3. Same statements of the series (2) and (3) were given by Lorentz [5], M. and Sh. Izumi [7] and R. Boas [6]. We shall state here a theorem generalizing the corresponding statements of these authors for k–fold Fourier series of type (1).

Definition. *Let* $f \in C(T^k)$, $\varphi :]0,1]^k \to [0,+\infty[$ *and* φ *be nondecreasing with respect to each coordinate of vector* **t**. *We say* $f \in Lip^{(m)}\ \varphi(t)$ *if there exists such a number* $A(m)>0$, *that*

$$\left\|\Delta_m f(x;t)\right\|_{C(T^k)} \le A(m)\varphi(t), \quad t \in]0,1]^k.$$

The following theorem is true

Theorem. *Let* $a_n \ge 0$, $n \in N_0^k$, *and function* φ *satisfies the following condition*

$$\int_{H_2(t,B)} \int_{H_1(t,B')} \varphi(x) \prod_{i\in B} x_i^{-m_i-1} \prod_{j\in B'} x_j^{-1} dx \le A_i(m)\varphi(t) \prod_{i\in B} t_i^{-m_i} \tag{4}$$

for each $B, B \subset M$, *where* $A_1(m) \in]0,+\infty[$. *Then* $f \in Lip^{(m)}\varphi(t)$ *if and only if*

$$\sum_{i=\left[\frac{n}{2}\right]}^{n} a_i \le A_2(m)\varphi\left(\frac{1}{n}\right),$$

where $A_2(m) \in]0,+\infty[$.

We note that the conditions (4) are multivariant analogues of the corresponding conditions of N. Bari and S.Stechkin [4].

I.Javakhishvili Tbilisi State University

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MATHEMATICS

S.Zotniashvili

On the Extension of Selfadjoint Operators from Hilbert Spaces to Frechet–Hilbert Ones

Presented by Corr. Member of the Academy N.Vakhania, July 18,1996

ABSTRACT. **The paper deals with symmetric and selfadjoint operators in Hilbert and Frechet–Hilbert spaces. Sufficient condition for such operations defined on Hilbert spaces to be extendible as operation of the same kind to Frechet–Hilbert spaces is given. Moreover the domains of the operators definition are essentially extended. In some cases the obtained operators are continuous although they were unbounded in the initial Hilbert spaces.**

The theory of selfadjoint operators is very important for the modern quantum mechanics and mathematical physics. The continuous selfadjoint operators in the Frechet–Hilbert spaces for the first time were defined and studied in [1–2] (the terminology is mainly used from [3]). Without requirement of continuity these operators were defined in [4] for the projective limits of sequences of Hilbert spaces. It should be noted that the latters contain the classes of Frechet–Hilbert, nuclear Frechet and countable–Hilbert spaces. In [5] the extension of Ritz's method for the equation with symmetric and positive definite operators was given.

Further we shall assume that the topology of the Frechet space is given by an increasing sequence of Hilbertian seminorms $\{\|\cdot\|_n\}$.This means that each $\|\cdot\|_n$ can be expressed by means of semiinner product $(.,.)_n$, i.e. for each $x\in E$ the equality holds $\|x\|_n = (x, x)_n^{1/2}$. Let $A:E\to E$ be a linear operator with a dense domain $D(A)$. If for $y\in E$ there exists an element $y^*\in E$ such that $(Ax,y)_n=(x,y^*)_n$ for $x\in D(A)$ and $n\in N$, then by setting $A^*y = y^*$ we can define the operator $A^*:E\to E$ which is called the Hilbertian adjoint operator for A. The operator A^* is different from the usual topological adjoint operator A'. The above operator A is called symmetric if

$$(Ax,y)_n=(x,Ay)_n$$

for any $x,y\in D(A)$ and $n\in N$. Note that for symmetric operator A we have $D(A)\subset D(A^*)$, i.e. the Hilbertian adjoint A^* is an extension of A.

A symmetric operator is called selfadjoint if $A=A^*$, i.e. $D(A)=D(A^*)$.

A Frechet–Hilbert space E is called a Frechet–Hilbert space (a H–Frechet space according the terminology of [1]) if its topology is generated by a sequence of

Hilbertian seminorms $\|x\|_n = (x,x)_n^{1/2}$ $(n \in N)$ and the space E complete in each seminorm topology. Therefore, in the case of Frechet–Hilbert space E is the quotient spaces $E/Ker\|.\|_n$ are Hilbert ones with the associated norms $\left\|\hat{k_n x}\right\|_n = \|x\|_n$, where $k_n{:}E \to E/Ker\|.\|_n$ are canonical mappings. Inner product on $E/Ker\|.\|_n$ is defined by equality $< k_n x, k_n y >_n = (x,y)_n$ and $<k_n x, k_n x>_n^{1/2} = \left\|\hat{k_n x}\right\|_n$ for each $x \in E$. Consequently, a Frechet–Hilbert space is the strict projective limit of a sequence of Hilbert spaces with respect to the canonical mappings

$$\pi_{nm}{:}(E / Ker\|\cdot\|_m, \hat{\|\cdot\|}_m) \to (E / Ker\|\cdot\|_n, \hat{\|\cdot\|}_n) \qquad (m \geq n).$$

This means that each $x \in E$ can be represented as a sequence $\{k_n x\}$, where $\pi_{nm}\, k_m x = k_n x\ (m \geq n)$.

Examples of Frechet–Hilbert spaces are: the space of all sequences $\omega = C^N$, the space $\omega x H$, where H is a Hilbert space, the Frechet spaces $(l^2)^N$ and $L^2_{loc}(x,\mu)$, where x is a locally compact space countable at infinity and μ is a Radon measure on x.

Now we give: a representation of Frechet–Hilbert space as the strict projective limit of a sequence of its complemented Hilbert subspaces; a representation of its strong dual space as the strict inductive limit of the same sequence of its complemented Hilbert subspaces being proved in [3] in case of real Frechet–Hilbert spaces.

Theorem 1. *Let E be a nonnormable Frechet space with an increasing sequence of seminorms generating the topology. Then the following statements are equivalent:*

a) E is a Frechet–Hilbert space with the sequence of seminorms $\|.\|_n\}$

b) The strong dual space $(E', \beta(E',E))$ *is the strict inductive limit of their increasing sequences of complemented subspaces* H_n *spanned by the polar* u_n^0 *of the neighborhood* $u_n = \{x \in E; \|x\|_n \leq 1\}$.

c) For each $n \in N$ *the space* E *is a topological sum of subspace* $Ker\|\cdot\|_n$ *and* $(H_n, \|\cdot\|_{nH_n})$, *where* $(H_n, \|\cdot\|_{nH_n})$ *is a Hilbert subspace of* E *with respect to the restriction* $\|\cdot\|_{nH_n}$ *of* $\|\cdot\|_n$ *to* H_n. *In particular* E *is the strict projective limit of the sequence* $\{(H_n, \|\cdot\|_{nH_n})\}$.

The proof of the theorem in case of complex Frechet spaces does not essentially differ from that given in [3], because H_n and H_n' considered as the subspaces of E and E' coincide with each other. Their norms are isometrical and inner products differ from each other as well as the inner products of complex Hilbert spaces and its dual.

From the Theorem 1 it follows that for each $n\in N$ the canonical mapping K_n is a projector of E onto H_n and its restriction on H_n is a topological isomorphism H_n onto $E/Ker\|\cdot\|_n$. Therefore each $x\in E$ can be represented as a sequence $\{\overset{(n)}{x}\}$, where $\overset{(n)}{x}$ is a projection of x in H_n which we call the trace of x in H_n. Hence, for each $h_1,h_2\in H_n$ the equalities are true

$$(h_1,h_2)=(h_1,h_2)_{n,H_n}=\langle k_n h_1,k_n h_2\rangle_n, \qquad (1)$$

where $(\cdot,\cdot)_{n,H_n}$ is a restriction of $(\cdot,\cdot)_n$ to H_n.

Let in notations of Theorem 1 $A:E\to E$ be a linear operator. Denote by A_n the projection of A on $E/Ker\|\cdot\|_n$ defined on $D(A_n)=K_n(D(A))$ by the equality

$$A_n(k_n x)=k_n(Ax) \qquad (2)$$

Denote also by $A^{(n)}$ the restriction of the operator A to H_n. If A is a symmetric operator. $D(A)\cap H_n$ is dense in H_n and H_n is an invariant subspace of A, i.e. $A(H_n)\subset H_n$, then according to (1) $A^{(n)}$ is a symmetric in H_n. It should be also noted that each symmetric operator in Hilbert space admits a sequence of invariant subspaces ([6], VII. 2, Lemma 2).

Theorem 2. *Let E be a Frechet–Hilbert space which is the strict projective limit of the sequence of its subspaces $\{(H_n, \|\cdot\|_{nH_n})\}$, $A:E\to E$ be a linear operator with a dense domain $D(A)$ and $A(H_n)\subset H_n$. Then the following statements hold:*

a) A is a symmetric operator in E if and only if A_n is a symmetric in a Hilbert space $(E/Ker\|\cdot\|_n, \hat{\|\cdot\|}_n)$ for each $n\in N$ with a dense domain $D(A_n)=K_n(D(A))$. Moreover, A is a continuous operator in $L_0(E)$ (i.e. $\|A_x\|_n\le C_n\|x\|_n$ for each $n\in N$ and $x\in E$) and is selfadjoint in E if and only if A_n is symmetric and $D(A_n)=H_n$.

b) For each $h\in D(A)\cap H_n$

$A_n(k_n h)=Ah=A^{(n)}h$, where A_n is defined by (2). Operator A is given on $D(A)$ by the equality $A\varphi=A\{\varphi_n\}=\{A_n\varphi_n\}=\{A^{(n)}\varphi_n\}$, where $\varphi=\{\varphi_n\}=\{k_n\varphi\}\in D(A)\subset E$.

Let now $(H,\|\cdot\|)$ be a Hilbert space, $\{H_n\}$ be an increasing sequence of their subspaces such that $\bigcup_{n\in N} H_n=F$ is a dense in H. Let also $j_{nm}:H_n\to H_m(n\leq m)$ $j_n:H_n\to F$ $(n\in N)$ be identical embeddings and $j^o_m j_{nm}=j_n$ $(n\leq m)$. If F is equipped with the inductive limit topology, then according to Theorem 1, the strong dual space $E=(F',\beta(F',F))$ is a Frechet–Hilbert space which can be represented as a strict projective limit of Hilbert spaces $\{H'_n\}$ $(H'_n=H_n)$ $(n\in N)$ sequence with respect to the adjoint mappings $j'_{nm}:H'_m\to H'_n$ $(n\leq m)$.

Theorem 3. *Let* $(H,\|\cdot\|)$ *be a Hilbert space,* $\{H_n\}$ *be an increasing sequence of their closed subspaces such that* $\bigcup_{n\in N} H_n=F$ *is dense in* H, A *be a symmetric operator in* H, $A(H_n)\subset H_n$ *and* $A^{(n)}$ *is a restriction of* A *on* H_n. *If for* $\varphi_m\in H_m$ *the equalities* $j'_{nm}A^{(m)}\varphi_m=A^{(n)}j'_{nm}\varphi_m$ $(n\leq m)$ *are hold, then*

$$\tilde{A}\varphi=\{A^{(n)}\varphi_n\},\ \varphi=\{\varphi_n\}$$

defines the symmetric operator on the Frechet–Hilbert space $E=(F',\beta(F',F))$.

For the illustration on Theorem 3 we apply it for the extension of the coordinate operator from the Hilbert space $L^2(R)$ onto Frechet–Hilbert space $L^2_{loc}(R)$. Let $H=L^2(R)$, then $H_n=L^2_o[-n,n]$ is the closed subspace of $L^2(R)$ of all functions which vanish almost everywhere outside $[-n,n]$; operators $j_{nm}:L^2_o[-n,n]\to$ $\to L^2_o[-m,m]$ $(n\leq m)$ and $j_m:L^2_o[-n,n]\to \bigcup_{n\in N} L^2_0[-n,n]=L^2_0(R)=F$ are identical embeddings and $Ax(t)=tx(t)$ is the coordinate operator defined in $L^2(R)$, $A^{(n)}$ is a restriction of the operator A on $L^2_o[-n,n]$. Obviously, $A(H_n)\subset H_n$ for each $n\in N$. Adjoint operators j'_{nm}: $L^2_o[-m,m]\to L^2_o[-n,n]$ and $j'_n:L^2_0(R)'=L^2_{loc}(R)\to$ $\to L^2_o[-n,n]$ are restrictions on $[-n,n]$. The equalities $j'_{nm}A^{(m)}\varphi_m=A^{(n)}j'_{nm}\varphi_m$ $(n\leq m)$ are valid and they show that the operations of restriction and multiplication by argument are commutative. Hence, by the equality $\tilde{A}\varphi(t)=\{A^{(n)}\varphi_n\}=t\varphi(t)$, $\varphi=\{\varphi_n\}\in L^2_{loc}(R)$ the extension of the operator A from the Hilbert space $L^2(R)$ onto the whole space $L^2_{loc}(R)$ is defined. Moreover according to the statement a) of the Theorem 2 we have $D(\tilde{A})=L^2_{loc}(R)$ and $\tilde{A}$ is a selfadjoint continuous operator belonging to the class $L_0(L^2_{loc}(R))$.

Theorem 3 also can be used for extension of selfadjoint differential operators L_∞, constructed in the papers [7–8].

N. Muskhelishvili Institute
of Computational Mathematics

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MATHEMATICS

T.Akhobadze

On Some Analogies of Riemann-Lebesgue Theorem for Functions of Multiple Variables-2

Presented by Academician L.Zhizhiashvili, March 4, 1996

ABSTRACT. **The present paper studies the limits in the restricted sense of the type**

$$\lim_{\lambda_1,\lambda_2\to+\infty}\iint\limits_I f(u,v)\alpha(\lambda_1 u,\lambda_2 v)\,dudv$$

for functions f and d from various classes.

In the recently published paper [1] for given function α and for function f from various classes we investigated the behaviour of the following limit in Pringaheim's sense

$$\lim_{\lambda_1,\lambda_2\to+\infty}\iint\limits_I f(u,v)\alpha(\lambda_1 u,\lambda_2 v)\,dudv,$$

where I is a two-dimensional interval.

The history of this problem comes from the well-known Riemann-Lebesgue theorem and further is developed in papers of various authors.

This paper aims to consider analogous problems for the limits in the bounded and restricted sense. From the obtained results particularly Kahane's [3] corresponding statements are implied with already known and new theorems for one-dimensional case.

As in the previous paper [1] for the presentation simplicity of obtained results we render them for two-dimensional case.

Denote by $L_p\{I\}$ $(1\le p\le+\infty)$ the class of Lebesgue integrable to the p-th power functions on two-dimensional I interval .If $p=+\infty$ we designate by $L_\infty\{I\}$ the class of essentially bounded function on I.

Theorem 1. *Let $\alpha\in L\{[0,+\infty)^2\}$. Then the necessary and sufficient condition for*

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_0^{+\infty}\int_0^{+\infty} f(u,v)\alpha(\lambda_1 u,\lambda_2 v)\,dudv$$

to exist in the bounded in the restricted sense for every function $f\in L_1\{[0,+\infty)^2]\}$ is presente of the limit

$$M(\alpha)\equiv\lim_{T_1,T_2\to+\infty}\frac{1}{T_1T_2}\int_0^{T_1}\int_0^{T_2}\alpha(u,v)\,dudv \tag{1}$$

also in the bounded in the restricted sense moreover this being the case

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_0^{+\infty}\int_0^{+\infty} f(u,v)\;\alpha(\lambda_1 u,\,\lambda_2 v)\,dudv=\left(\int_0^{+\infty}\int_0^{+\infty} f(u,v)\,dudv\right)M(\alpha).$$

The analogous statement is valid (see [1], corollary 1) for the limits

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_{-\infty}^{+\infty}\int_{-\infty}^{+\infty} f(u,v)\ \alpha(\lambda_1 u,\lambda_2 v)dudv$$

both in the bounded and in the restricted sense.

Definition 1. *Let γ be non- negative locally-integrable function on $[0,+\infty)^2$. We shall consider mappings $(T_1,T_2)\to\varphi_i(T_1,T_2)$, where $T_1,T_2\geq 1$ and $\varphi_i(T_1,T_2)>0$ $(i=1,2)$. We say that a function γ belongs to the class M_1 $(\gamma\in M_1)$ if for every bounded functions φ_1 and φ_2 (under the conditions $\lim\limits_{T_1\to+\infty}\varphi_1(T_1,T_2)=0$, $\lim\limits_{T_2\to+\infty}\varphi_2(T_1,T_2=0)$ uniformly concerning bounded T_2 and T_1 respectively)*

$$\text{a)}\quad \lim_{T_1+T_2\to+\infty}\frac{1}{T_1T_2}\int_0^{\varphi_1(T_1T_2)}\int_0^{T_2}\gamma\ (u,v)dudv=0$$

and

$$\text{b)}\quad \lim_{T_1+T_2\to+\infty}\frac{1}{T_1T_2}\int_0^{T_1}\int_0^{\varphi_2(T_1T_2)}\gamma(u,v)dudv=0$$

Theorem 2. *Let α be a logically-integrable to the p-th $(1\leq p\leq+\infty)$ power function; moreover suppose that $|\alpha|\in M_1$. Then in order the limit*

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_a^b\int_c^d f(u,v)\,\alpha(\lambda_1 u,\lambda_2 v)\,dudv\quad (0\leq a<b<+\infty,\ 0\leq c<d<+\infty)\tag{2}$$

should exist in the restricted sense for every function $f\in L_q.\{[a,b]\times[c,d]\}$, where $\frac{1}{p}+\frac{1}{q}=1$ (if $p=1$, assume that $p=\infty$), it is necessary and sufficient that

$$1)\ \frac{1}{T_1T_2}\int_{\frac{a}{b}T_1}^{T_1}\int_{\frac{c}{d}T_2}^{T_2}\left|\alpha(u,v)\right|^p dudv$$

be bounded as $T_1,T_2\geq 1$; 2) should exist limit (1) in the restricted sense. The conditions 1) and 2) should also be in the restricted sense

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_a^b\int_c^d f(u,v)\,\alpha(\lambda_1 u,\lambda_2 v)\,dudv=\left(\int_a^b\int_c^d f(u,v)dudv\right)M(\alpha)\tag{3}$$

is valid.

Remark 1. Formulated conditions 1) and without assumption $|\alpha|\in M_1$ are sufficient for the existence of the limit (2).

Remark 2. If in the Theorem 2 $a=c=0$ then the condition $|\alpha|\in M_1$ is unnecessary. If a or c are positive then the condition $|\alpha|\in M_1$ can't be weakened.

Definition 2. Suppose that γ is non-negative, locally-inregrable function on $[0,+\infty)^2$ and let $T\to\varphi_i(T)$ ($\varphi_i(T)>0$, $T\geq 1$, $i=1,2$) be homogeneous functions. We say that γ belongs to the class M_2 ($\gamma\in M_2$) if for every φ_1 and φ_2

$$\text{a)} \lim_{\substack{T\to+\infty \\ \varphi_1(T)\to 0}} \frac{1}{T^2}\int_0^{\varphi_1(T)}\int_0^T \gamma(u,v)\,du\,dv = 0,$$

$$\text{b)} \lim_{\substack{T\to+\infty \\ \varphi_2(T)\to 0}} \frac{1}{T^2}\int_0^{T}\int_0^{\varphi_2(T)} \gamma(u,v)\,du\,dv = 0,$$

Condition a) means that for every $\varepsilon>0$ there exists such positive number N that if $T>N$ and $\varphi_1(T)<\dfrac{1}{N}$ then

$$\frac{1}{T^2}\int_0^{\varphi_1(T)}\int_0^T \gamma(u.v)\,du\,dv < \varepsilon.$$

Analogously condition b) can be defined.

Theorem 3. *Let α be a locally-integrable to the p-th ($1\leq p<\infty$) power function. Suppose also that $|\alpha|\in M_2$. Then the necessary and sufficient condition limit* (2) *to exist in the bounded sense for every function $f\in L_q\{[a,b]\times[c,d]\}$, $\dfrac{1}{p}+\dfrac{1}{q}=1$ (if $p=1$ assume that $q=\infty$), is that*: 1) *for every number β, $\beta\geq 1$, should exist such positive number M that for all $T_1,T_2\geq 1$, $\dfrac{1}{\beta}\leq\dfrac{T_1}{T_2}\leq\beta$, the following estimation*

$$\frac{1}{T_1T_2}\int_{\frac{a}{b}T_1}^{T_1}\int_{\frac{c}{d}T_2}^{T_2} |\alpha(u,v)|^p\,du\,dv \leq M$$

takes place; 2) *should exist limit* (1) *in the bounded sense. If the conditions* 1) *and* 2) *are fulfilled then* (3) *is valid in the bounded sense.*

Remark 3. Formulated conditions 1) and 2) in Theorem 3 without assumption $|\alpha|\in M_2$ are sufficient for limit (2) existence in the bounded sense.

Remark 4. If in the Theorem 3 $a=c=0$ then the condition $|\alpha|\in M_2$ is unnecessary; if a or c are positives then the condition $|\alpha|\in M_2$ in the definite sense can't be weakened.

I.Javakhishvili Tbilisi State University

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MATHEMATICS

E.Nadaraya, R.Absava, L.Nadaraya, M.Vezirishvili

Testing the Equality of Two Regression Curves

Presented by Academician G.Chogoshvili, February 19, 1996

ABSTRACT. **Criteria for checking the hypothesis about the equality of regression function have been constructed from two nonparametric models.**

Let two sets of observations be given, the first one formed by the observations:

$$Y_j^{(1)} = \mu_1(x_j) + \varepsilon_{1j},\ j = \overline{0,2n},$$

and the second formed by

$$Y_j^{(2)} = \mu_2(x_j) + \varepsilon_{2j},\ j = \overline{0,2n},$$

where $\mu_1(x)$ and $\mu_2(x)$, $x\in[-\pi,\pi]$ are unknown regression functions. Numbers x_i are defined by $x_i = -\pi + \dfrac{2\pi i}{2n+1}$, $i = \overline{0,2n}$, totalities of errors ε_{1j} and ε_{2j}, $j = \overline{0,2n}$ are independent. Within these totalities the errors are independent and identically distributed with zero means and variances σ_1^2 and σ_2^2, respectively, satisfying conditions $E\varepsilon_{11}^4 < \infty$ and $E\varepsilon_{21}^4 < \infty$.

Suppose that $\mu_i(x)$, i = $\overline{1,2}$ is representable as a convergent in $L_2(-\pi,\pi)$ series with respect to orthonormal trigonometric system

$$\{(2\pi)^{-1/2},\ \pi^{-1/2}\cos ix,\ \pi^{-1/2}\sin ix\}.$$

Let's consider the estimator [1] of the regression function $\mu_i(x)$, constructed by means of Chentsov's [2] projection method of the form

$$\mu_{nN}^{(i)}(x) = \frac{a_{0n}^{(i)}}{2}+\sum_{k=1}^{N} a_{kn}^{(i)} \cos kx+b_{kn}^{(i)}\sin kx,\ i = 1,2,$$

where $N = N(n)\to\infty$ as $n\to\infty$ and

$$a_{kn}^{(i)} = \frac{1}{\pi}\sum_{j=0}^{2n} Y_j^{(i)}\,\Delta_j\cos kx_j\quad b_{kn}^{(i)} = \frac{1}{\pi}\sum_{j=0}^{2n} Y_j^{(i)}\,\Delta_j\sin kx_j,$$

$\Delta_j = x_j - x_{j-1}$.

In practice it is often necessary to test the equality of regression functions i.e. we have to test hypothesis H_0: $\mu_1(x) = \mu_2(x)$ for all $x\in[-\pi,\ \pi]$ against the alternative H_1:$\mu_1(x)\neq\mu_2(x)$, $x\in[-\pi,\pi]$.

As the distance between $\mu_{nN}^{(1)}(x)$ and $\mu_{nN}^{(2)}(x)$ we consider the statistic

$$V_{nN}= \frac{2n+1}{2\pi(2N+1)}\int_{-\pi}^{\pi}\left[\mu_{nN}^{(1)}(x)-\mu_{nN}^{(2)}(x)\right]^2 dx,$$

that enables to construct the criterion of the given level for testing the null hypothesis H_0. With this purpose in the given paper the limiting distribution of the functional V_{nN}

under the null hypothesis is obtained; the consistency of the constructed criterion is established, as well as the local behavior of the power of the same criterion against the sequence of close alternatives to the hypothesis H_0 is stated.

Theorem 1. *Let* $\frac{N^2 \ln N}{n} \to 0$ *as* $n\to\infty$ *. Then given null hypothesis* H_0 *is valid, we have*

$$\sqrt{N+\frac{1}{2}}\,(V_{nN}-\sigma^2)\sigma^{-2} \xrightarrow{d} \xi,\ \sigma^2=\sigma_1^2+\sigma_2^2.$$

The symbol " $\xrightarrow{d}$ *" denotes the weak convergence,* ξ *is a random variable normally distributed with zero mean and unit variance.*

Theorem 1 enables us to test the hypothesis H_0*, according to which* $\mu_1(x)=\mu_2(x)$ *for all* $x\in[-\pi,\pi]$.

Let σ^2 *be known. The critical region is established approximately by the inequality*

$$V_{nN}\ge d_{nN}(\alpha), \tag{1}$$

where

$$d_{nN}(\alpha)=\sigma^2(1+\sqrt{2}\,(2N+1)^{-1/2}\lambda\alpha),$$

and λ_α *is* α*-tl quantile of standard normal distribution.*

Let σ^2 be unknown. Denote $Z_j=Y_j^{(1)}-Y_j^{(2)}$ and as an estimator of σ^2let's introduce the statistic

$$S_n^2=\frac{1}{4n}\sum_{j=1}^{2n}(Z_j-Z_{j-1})^2.$$

S_n^2 is an unbiased and $\sqrt{N}$ is consistent estimator of the variance σ^2.

Corollary of Theorem 1. *Let the conditions of Theorem 1 are valid. Then*

$$\sqrt{N+\frac{1}{2}}\,S_n^{-2}(V_{nN}-S_n^2)\xrightarrow{d}\xi.$$

In this case the critical region is established by the inequality $V_{nN}\ge\tilde{d}_{nN}(\alpha)$*, where* $\tilde{d}_{nN}(\alpha)$ *is obtained from* $d_{nN}(\alpha)$ *substituting* S_n^2 *for* σ^2.

Studying the asymptotic properties of the test (1) (i.e. the power function behaviour as $n\to\infty$*), let's first consider the question whether the test is consistent.*

Theorem 2. *Let* $\frac{N^2\ln N}{n}\to 0$*, as* $n\to\infty$*. Then* $P_{H_1}\{V_{nN}\ge d_{nN}(\alpha)\}\to 1$ *as* $n\to\infty$*, i.e. the test defined by (1) is consistent against any alternative* H_1*:* $=\mu_1(x)\ne\mu_2(x)$,

$$\int_{-\pi}^{\pi}(\mu_1(x)-\mu_2(x))^2\,dx>0.$$

Thus the power of test (1) under any fixed alternative tends to 1 as $n\to\infty$. However deeper properties of the test are revealed when studying the test reactions on "small" deviations from the hypothesis under test (i.e. when instead of fixed alternative the sequence of alternatives approaching to the main hypothesis H_0 as $n\to\infty$ is considered).

In our case the sequence of close alternatives has the form

$$H_{1n}: \nu_n(x)=\mu_1(x)-\mu_2(x)=\nu(x)\gamma_n,\ \nu(x)\ne 0, \tag{2}$$

when $\gamma_n \to 0$ in an appropriate.

Theorem 3. *Let* $v(x)$, $x\in[-\pi,\pi]$ *be bounded and have bounded derivative. If* $N = 2n^{\delta}$, $\gamma_n = n^{-1/2+\delta/4}$, $0<\delta<\frac{1}{2}$, *then the statistic* $(2N+1)^{1/2}(V_{nN}-\sigma^2)$ *under the alternative* H_{1n} *has the limiting normal distribution* $\left(\frac{1}{2\pi}\int_{-\pi}^{\pi} v^2(x)dx, \sqrt{2}\sigma^2\right)$.

Note. It follows from Theorem 3 that closer alternatives of the form (2) (i.e. under $\gamma_n \cdot n^{1/2-\delta/4} \to 0$) are not distinguished from H_0 by this test (i.e. $P_{H_{1n}}(V_{nN} \ge d_{nN}(\alpha)) \to \alpha$), and for more remote alternatives (i.e. under $\gamma_n \cdot n^{1/2-\delta/4} \to \infty$) it preserves the consistency property (i.e. $P_{H_{1n}}(V_{nN} \ge d_{nN}(\alpha)) \to 1$).

Thus the local behaviour of power $P_{H_{1n}}\{V_{nN} \ge d_{nN}(\alpha)\}$ is the following:

$$P_{H_{1n}}(V_{nN} \ge d_{nN}(\alpha)) \to 1-\Phi\left(\lambda_\alpha - \frac{1}{\sqrt{2}\sigma^2}\frac{1}{2\pi}\int_{-\pi}^{\pi} v^2(x)dx\right), \qquad (3)$$

where $\Phi(u) = \frac{1}{\sqrt{2\pi}}\int_{-\infty}^{u} \exp\left\{-\frac{x^2}{2}\right\} dx.$

Since $\int_{-\pi}^{\pi} v^2(x)dx > 0$ and is equal to zero if and only if $v(x) = 0$, it follows from (3) that the test for treating the hypothesis $H_0: \mu_1(x) \equiv \mu_2(x)$ against an alternative of the form (2) is asymptotically strictly unbiased.

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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CYBERNETICS

G.Badiashvili

Fuzzy-Conceptual Method of Knowledge Presentation and Acquisition in the Medical Intelligence System

Presented by Academician V.Chavchanidze, July 18, 1996

ABSTRACT. **Formalism of conceptual systems and general theory of management becomes a basis of conceptual idea of knowledge presentation. Different methods of concept formation are also based on this idea, i.e. analytic-heuristic method, analytic-filtration method, etc. It must be mentioned, that this approach completely describes the problems of medical diagnostics and prognostics. Simultaneously together with the theory of conceptual systems the theory of fuzzy sets is based on the concept of "linguistic variable". This theory enables us to describe those processes, about which we have no complete information, or this information is not exact. At present this theory is under the development.**

The method, described in the article is based on the principles of these two theories and at certain step their linkage is made. This algorithm makes a modelling of decision-making by expert and provides the acquisition of machine knowledge in given problem field. At each step decision is made on the basis of early experience. It must be also mentioned, that the described system needs minimal volume of machine memory for acquisition the experience.

Let's assume that we have any A object (in this case any nosological unit, i.e. acute sepsis). It is characterized with definite $S=\{S_1,S_2,\dots,S_m\}$ set of sings. For S set we have the set of values $Q=\{q_0, q_1, q_2, \dots, q_N\}$. We can imply that these values are given in the form of linguistic variables (Fig.1). Let's also assume, that q_0 and q_N express the absence or highest possible presence, i.e. "not expressed" and "expressed at most". On the basis of carried out observations (in our case the study of medical records) and the knowledge, obtained from experts for each S_m m = 1, ..., M we can define the $P_m(q_i)$ probability of q_i event. From $\{q_n\}$ n = 0, ..., N we separate the q_{i_1}, q_{i_2}, ..., $q_{i_{n_m}}$ subsequency. For this subsequency $P_m(q_{ij}) > 0$, $j = 1,\dots, n_m$ and $P_m(q_{i_1}) + P_m(q_{i_2}) + \dots + P_m(q_{i_{n_m}}) = 1$ (Fig.2).

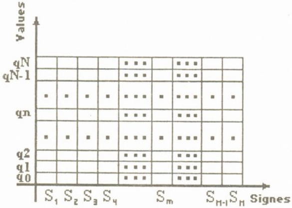

Fig.1.

Informational unit is characterized with four properties (**object**, **sign**, **value**, **reliability**) [9]. The function corresponds to the sign which gives the values of object

$$\text{value} = \text{sign (object)} \qquad (*)$$

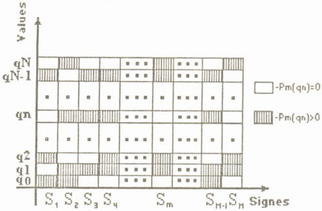

Fig.2.

Values of **reliability** define the level of information vagueness. For its evaluation we use two qualifications: "possibility" and "necessity". Measure of vagueness evaluates the level of reliability of the subject, considering the current level of informing towards A event. If S is a universal set and $\varnothing$ empty set, then we have an axioms:

$$g(\varnothing) = 0 \text{ and } g(S) = 1 \qquad (1)$$

for practical problem we have a weaker axioms:

$$A \subseteq B \Rightarrow g(A) \leq g(B) \qquad (2)$$

This axiom mean as follows: if A event is followed by other B event, then there always exists at least the same reliability of B event, as in the case of A event.

According to the results, obtained in [9], allocation $\pi(q_i)$ and membership function $\mu_{Sm}(q)$ will be given by the following formula:

$$\pi_m(q_i) = \sum_{j=1}^{n_m} \min\left(P_m(q_i), P_m(q_j)\right), \qquad (3)$$

and

$$\mu_{Sm}(q) = \sum_{q':P_m(q')\leq P_m(q)} P_m(q') \qquad (4)$$

Limited values of $g(A)$ vagueness measure are measures of "possibility" $\Pi(S_m)$ and of "necessity" $N(S_m)$, $m = 1,\dots, M$. These measures are calculated by (3) and (4) as:

$$\Pi(S_m) = \sup_{q\in Q} \min\left(\mu_{Sm}(q), \pi_m(q)\right) \qquad (5)$$

and

$$N(S_m) = \inf_{q\in Q} \max\left(\mu_{Sm}(q), 1 - \pi_m(q)\right) \qquad (6)$$

It is also very important to show the following relations between $\Pi(Sm)$ and $N(Sm)$:

$$N(Sm_1 \cap Sm_2) = \min(N(Sm_1), N(Sm_2)) \qquad (7)$$

$$\Pi(Sm_1 \cup Sm_2) = \max\left(\Pi(Sm_1), \Pi(Sm_2)\right) \qquad (8)$$

$$(N(S') = 1 - \Pi(\bar{S}')) \Leftrightarrow 1 - (N(S')), \quad \forall S' \subseteq S \qquad (9)$$

where $\bar{S}'$ is an addition of S' subsequency up to S set.

(7)-(9) formulae enable us to make final evaluation of diagnosis according to the fact whether we have the union or intersection of S set.

It is necessary to show two following relations between the measures of "possibility" and "necessity":

$$\forall S' \subseteq S, \text{ if } \Pi(S') = 1,\ N(S') > 0 \text{ and if } \Pi(S') < 1,\ N(S') = 0 \qquad (**)$$

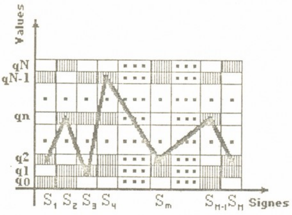

Fig.3.

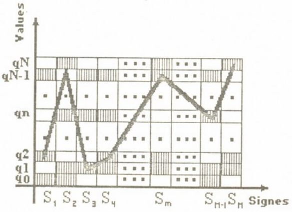

Fig.4.

On purpose to identify A object, on each next step of working the system will have to make a "trajectory", crossing the corresponding values of $S1,\ldots,Sm$ signs. In this case we may have two occasions: this "trajectory" may cross (Fig.3) the subsequancy of values change or not (Fig.4).

In the first case the identification of A object is not a problem. Before making next step, system will "reallocate" the probabilities in assumable values on the basis of input information. Old values of possibilities will be removed and new values will be saved.

In the second case the system directly diagnoses that the object to be identified is not A (in this case $\Pi(S') < 1, \forall S' \subseteq S$. If this result doesn't cause suspicion of the user, we can pass to another problem. If it causes any suspicion, then it is necessary to involve an expert and if the expert confirms that the object belongs to class A, then the connections between the S set elements (logical connections) must be changed and new subsequency of assumed values must be identified on the basis of input information. Old values of probabilities will be removed and new values saved.

I.Javakhishvili Tbilisi State University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

PHYSICS

T.Chelidze, T.Berikashvili, T.Peradze, I.Stamateli, L.Chachkhiani

Influence of Plastic Deformation on the Grain Substructure of Ti-Ta Alloys

Presented by Academician G.Tsintsadze, May 14, 1996

ABSTRACT. **Influence of plastic deformation on quenching-induced phase transformation and inner crystal structure of Ti-Ta alloys has been studied by means of electronic microscopy and microdiffraction methods.**

The role of these transformations and structures in form memory effect has been fixed.

The scope of the scientific works [1,2] carried out by the authors comprises investigation of Ti-Ta alloys (Ta content in the range of 5-85%). The main interest of the group was centered on such phenomena as form memory effect and super plasticity which had been detected for the first time in deformed alloys of mentioned composition. *X*-ray analysis, investigation of plastic modulus dependence on Ta concentration, differential thermal analysis of quenched samples of mentioned composition after annealing at 1000°C (within 75 hours) enabled to identify types of metastable structures in alloys and their dependence on the mentioned effect. Thus in Ta 28-50% alloys quenching fixes existence of α'' orthorombic phase only, whereas Ti-50-68% Ta and Ti-68-85% Ta are formed in two-phase $(\alpha''+\beta_{met})$ and β_{met} single phase states (β_{met}- volume centered cubic lattice) receptively.

Concentration of metastable state versus temperature diagram has been plotted. Mechanical tests and dilatometric experiments detected that form memory effect is manifested in α'' and $(\alpha''+\beta_{met})$ quenched alloys, while superplasticity is characteristic for all structures mentioned above. Role of the β_{met} phase in both effects consists of its participation in deformation thermoplastic martensit $\beta \rightleftarrows \alpha''_{def}$ transformation caused by plastic deformation of alloys. In alloys with single phase α'' structure deformation does not cause any phase transition. This fact has been proved by *X*-ray analysis. Deformation affects only the inner "fine" structure of grains. Form memory effect in such structures points to their thermoreversibility.

Present paper deals with indentification of the "fine" structure. Applied method: differential electronic microscopy, electronic microdiffraction (JEOL JEM - 200A type electronic microscope). Typical results are presented by data obtained for Ti-48.49% Ta and Ti-64.20% Ta alloys. The first alloy is very close to equilibrium limit of $\alpha''/(\alpha''+\beta_{met})$ metastable phases, i.e. it corresponds to maximum concentration of Ta. In this case α'' could be fixed by quenching. As for Ti-64.20% Ta it is of maximum concentration as well. (Near the $(\alpha''+\beta_{met})/\beta_{met}$ boundary). Increase of Ta content lowers temperature of martensit $\beta \rightleftarrows \alpha''$ transformation [3]. Thermally initiated phase transformation (which is able to affect the form memory effect) was avoided by plastic deformation of alloys at temperature below T_0 (T_0 max temperature of the phase existence). Control of the phase structure was accomplished by *X*-ray diffraction. For example, Ti-48.49% Ta (T~410°C) alloys were deformed at ~200°C, while Ti-64.20% Ta (T_0 = 140°C) has been treated at room temperature.

Figure 1(a) presents electro-microscopic images obtained for structures created by plastic deformation of grains in α'' orthorombic Ti-48.45% Ta alloys. Structure is lamellar type with differently oriented parallel plates. Analysis of reflexes on corresponding electro-micro-diffraction image (Fig.1 b) is given by scheme (Fig. 1c). It is evident that <112> zone is parallel to electron beam whereas the image represents superposition of section of three {112} inverse lattices. Issued from the symmetry of reflex location it is evident that each of the mentioned three sections makes (two by two) twin structures with others (Fig. 1c). Plate directions A_1A_2 and B_1B_2 (see Fig. 1 d) presented on the scheme could be parallel to the traces of {113} and {112} planes. Thus, lamellar structures shown in Fig.1 are presented by microtwin systems filling the whole intergrane volume. Twin orientation planes could coincide with {113} and {112} planes. $(\bar{1}\,\bar{1}\,3)$ $(\bar{1}\,3\,\bar{1})$ (see Fig.1 d) planes are perpendicular to microscopic image plane, whereas $(\bar{1}\,\bar{1}\,2)$ and $(\bar{1}\,2\,\bar{1})$ make 80 angle with the latter.

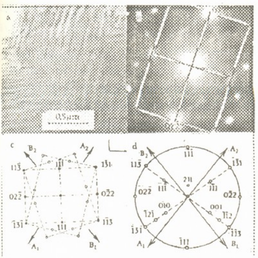

Fig.1. Ti-48.49%Ta alloys. (a) - micro-twin structure; (b) - microdiffraction of the structure; (c) - scheme obtained basing on the analyses of microdiffractions; (d) - stereography of α" orthorombic phase inverse lattice.

Schemes of expected diffraction images for {112} and {113} twin orientation planes in orthorombic systems have been calculated and plotted basing on [4] methods. Comparison of experimental micro-diffraction images with calculated ones revealed that {113} is a twin plane. Hence plastic deformation on quenched single phase α'' Ti-Ta alloys (sliding along {113} plane mechanism) creates definitely oriented intergrane twin structures.

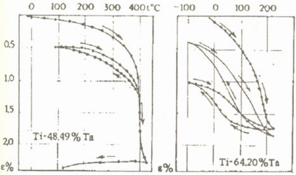

Fig.2.

ε - t dependence of form memory effect manifestation.

(a) - Ti-64.20%Ta alloys (ε_0=4.90%, ε_0= 7.95%);

(b) - Ti-48.49%Ta alloys (ε_0 = 3.95%).

Form memory effect in alloys under consideration is defined by deformation structure and displays thermal reversibility of the latter. In contrast to the alloys with (α''+β_{met}) initial structure, "velocity" of form memory effect in α''-single phase plastically deformed structures is high (Fig. 2b). The process is "explosion" like and takes place within the narrow temperature interval (~10-15°C). Fig. 2b testifies that at the initial period of form restoration the microtwin system is thermoplastic (thermoreversible) whereas further on it undergoes through relaxation.

In $(\alpha''+\beta_{met})$ structures form memory effect conditioned by plastic deformation process occurs at rather high temperatures and remains reversible within the whole temperature range (heating and cooling rate amounted to 0.17 K sec^{-1}). In Ti 64.20% Ta alloy this reversible process is maximal (Fig. 2a), i.e. in Ti-64.20% Ta alloys at the edge of $(\alpha''+\beta_{met})$ deformation structure responsible for the effect is max. developed. The structure is shown in Fig. 3 (a,b). Two types of grain structure is evident. In one case (see Fig. 3 a,b) image and corresponding micro-diffraction are similar to that of α''-phase deformed microtwin structure. It is obvious that the mentioned structure originates from plastic deformation of α'' phase in the initial $(\alpha''+\beta_{met})$ structure. Double oriented plane is {113} type.

Images of grains with developed dislocation structure correspond to α''_{def} grain structure obtained by $\beta_{met} \rightleftarrows \alpha''_{def}$ deformation martensit transition. Direction of dislocation image coincides with <113> direction.

Thus, in Ti-Ta alloys with plastically deformed initial $(\alpha''+\beta_{met})$ structure two mechanisms of form memory effect are possible. The first one consists of reversible relaxation of microtwin structures created by plastic deformation of α'' phase grains. The second - corresponds to thermoelastic martensit $\beta_{met} \rightleftarrows \alpha''_{def}$ transformation. Obtained data enable to explain results presented in [5]. According to the mentioned work form memory effect in $(\alpha''+\beta_{met})$ structures is inversely proportional to quantity of β_{met} phase. In case of $\beta_{met} \rightleftarrows \alpha''_{def}$ transformation both dislocation movement and process of developed dislocation system formation are irreversible. On the other hand quantity of the α'' phase decreases and correspondingly lowers the share of this phase in cumulative manifestation of form memory effect.

For all types of Ti-Ta alloys subjected to deformation above ~ 4-45% rate of the form memory effect is independent from deformation level. In alloys with $(\alpha''+\beta_{met})$ initial structure $\beta_{met} \rightleftarrows \alpha''_{def}$ deformation martensit transformation occurs only in case of deformation above ~1.5%. So, within the mentioned deformation area formation of deformed structure is completed. Thus along with intensification of

Fig.3 Electronmicroscopic images of grains in deformed Ti-64.20%Ta alloys.
(a) - deformed α''-phase structure;
(b) - structure of α''_{def} phase grains obtained by $\beta \rightarrow \alpha''$ deformation martensit transformation;
(c)- microdiffraction of deformed α''_{def} phase.

reversible form memory process in deformed ($\alpha''+\beta_{met}$) initial structure, increasing of β_{met} phase quantity in quenched alloys takes place. Results obtained for the inner structure of deformed α''_{def} phase crystals enables to conclude that displacement of dislocations in case of the thermally induced form memory effect results in both slackening of the mentioned effect and preservation of arose strain which in its turn prevents from the thorough relaxation of the structure. The strain is accumulated in grains and while cooling of samples (after heating up to the certain temperature) causes inverse process - restoration of initial plastically deformed structure ($\alpha''_{def} \rightarrow \beta_{met}$). It should be mentioned that $\beta \rightleftarrows \alpha''_{def}$ transformation and microtwin structure formation in α'' phase results in creation of fluidity plato on deformation curve.

Obtained data helps to conclude that: plastic deformation of quenched α'' and ($\alpha''+\beta_{met}$) phase structure Ti-Ta alloys results in creation of developed thermoelastic microtwin system with the α'' orthorombic phase. The twins are presented by parallel planes with various orientation (<113>, <110>, <112>). Fixed twin plane orientation {113} is not typical for orthorombic systems. Deformation of β_{met} phase initiates $\beta_{met} \rightleftarrows \alpha''_{def}$ thermoelastic (reversible) martensit transformation which is accompanied by formation of developed dislocation structure. The mentioned structure exerts negative influence on magnitude of the form memory effect in alloys but, at the same time, intensifies reversibility of the process. *X*-ray diffraction, electromicroscopy and microdiffraction enables to assert that any other types of structure transformation or alternation of inner structure of grains that not detected. Hence, form memory effect and manifestation of super elasticity in Ti-Ta alloys are conditioned by formation of the above mentioned structures only.

Georgian Technical University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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PHYSICS

V.Kashia, P.Kervalishvili, Academician R.Salukvadze

Study of Characteristics of the High-Temperature Thermo-Emissive Converters with Tungsten Electrodes

Presented June 21, 1996

ABSTRACT. **Experimental results of study of heat-energetic characteristics of the cylindrical cesium-plasma thermo-emissive converter (TEC) are presented. A tungsten layer with (110) preferable orientation served as an emitter. The role of collector was performed by polycrystalline or prepared by emitter's technology tungsten layer. The emitter's temperature under TEC generating conditions was T_E=2000-2200 K, and the collector temperature - T_C=1000-1300 K.**

It is shown that the use of the tungsten layer of (110) preferable orientation as a collector in the TEC with tungsten emitter is highly promissing but it requires high level of manufacturing - exploitation of converter units to prevent degradation and maintaining long-term stability of the device initial energetic parameters which at an average by 25 % prevail over parameters of the converter with polycrystalline tungsten collector.

At present practically in all cesium-plasma thermo-emissive converters (TEC) the emitter role is performed by polyfacet tungsten monocrystalline layer with (110) – preferable orientation plated by method of chemical transfer in chloride area upon the molibden monocrystal of [111] direction. Such emitters at rather high temperatures (2000 - 2200 K) are able to function stably for tens of thousands of hours [1,2].

Under the generation conditions the growth of the TEC emitter's temperature causes the considerable increase of device efficiency, but at the same time the emitter material transfer within the gap between electrodes becomes more intensive. This process already at the stage of TEC electrodes' treatment provokes the collector surface coating with various compositions of emitter material. It is ascertained that the collector material structure determines the emissive properties and elementary composition of the adsorption layer formed upon the collector's surface. All this in total has a decisive influence on the TEC parameters stability and technical recourse of the device. Many questions concerning this problem are solved while the TEC electrodes are manufactured from the same material e.g. tungsten [2].

Just because of this the high practical importance has the study of cesium – plasma TEC volt-amperic characteristics within a high temperatures range for tungsten electrodes prepared by different technologies. Such data today practically are not available whereas they are very necessary for calculating and forecasting parameters of thermo-emissive one- and multielement electrogenerating channels and nuclear power plants. Though it is done for one fixed temperature of the polycrystalline tungsten collector [3], the total block of volt-amperic characteristics for various densities of a generated current do not exist.

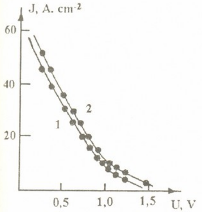

Fig.1. Curves of TEC's totally optimized voltamperic characteristics under T_E = 2200 and T_C = 1300 K for various collectors: 1 - polycrystalline tungsten; 2 - monocrystalline (110) tungsten.

There are presented in this work results of the experimental study of volt-amperic characteristics of the cylindrical model cesium-plasma TECs at the emitter and collector temperatures T_E=2000-2200 K and T_C = 1000-1300 K respectively.

In both TECs the emitter's role was performed by the tungsten layer of 200 mcm in thickness plated by means of the method of chemical transfer in a chlorides area upon the molibden monocrystal with (110)-preferable orientation. In the first converter a polycrystalline tungsten layer of 30 mcm in thickness plated upon the niobium-zirconium alloy served as a collector and in the second converter – a tungsten monocrystalline layer with (110)-preferable orientation produced by analogous to emitter technology. The gap between electrodes at the room temperature was for the first device – 0.25 mm and for the second one – 0.27 mm.

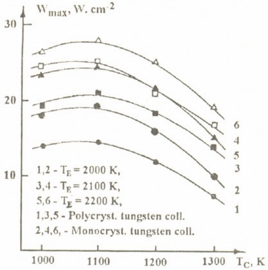

Fig.2. Wmax = f (Tc) dependences of TEC for various collectors and under emitter's different temperatures: 1,2 - T_E = 2000; 3,4 - 2100; 5,6 - 2200 K. 1,3,5 - policrystalline tungsten collector; 2,4,6 - monocrystalline tungsten collector.

Experiments were carried out on the electrovacuum testing bench [3] in which a vacuumization of TEC working volume was obtained by means of non-oil pumps and after gas evacuation from an electrodes area the static vacuum in the volume was $1 \cdot 10^{-4}$ Pa.

The time-temperature conditions of gas evacuation from electrodes, working volume and other units of the devices as well as methods for studying-processing volt-amperic characteristics are described in [1,3]. Analysis of results shown that electric parameters of various devices at electrodes' equal temperatures differ considerably from each other.

In the Fig.1 devices optimazed volt-amperic characteristics are presented from which it is evident that the curves of volt-amperic characteristics of the TEC with the monocrystalline tungsten collector is shifted towards high voltages almost parallel to curves of volt-amperic characteristics of the TEC with the polycrystalline tungsten collector.

In the Fig.2 drawn on the base of volt-amperic curves diagrams are presented which show the dependence of the devices maximum power on collector's various temperatures. Curves 1, 3, 5 correspond to the TEC with polycrystalline tungsten collector and curves 2, 4, 6 – to the TEC with monocrystalline tungsten collector.

It is evident from the Figure that the polycrystalline as well as monocrystalline tungsten collectors' temperature is T_C=1100 K but the power of the TEC with monocrystalline tungsten collector is by 25 % more than in case of polycrystalline collector (Fig.3). An analogous phenomenon is observed almost for all characteristics.

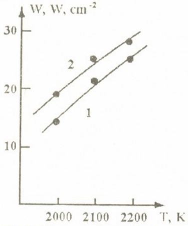

Fig.3. W_{max} = f (T_E) dependence under T_C = 1100 K for various collectors: 1 - polycrystalline tungsten; 2 - monocrystalline tungsten.

In the TEC with polycrystalline tungsten collector cesium's optimum pressure is approximately two times as big as that of the TEC with monocrystalline collector at the equal temperatures of emitters.

It corresponds to the thermo-emissive conversion theory [4] in accordance with which the more is the electrodes exit work, the less is cesium pressure needed for obtaining a minimum value of the exit work of the electrode active surface. The cesium-plasma TEC initial characteristics considerably prevail over parameters of the TEC with polycrystalline tungsten collector. However after functioning of the TEC with monocrystalline tungsten collector during 100 hours in generating regime the degradation of device's parameters was observed as well as their stabilization on the level of output parameters of the converter with polycrystalline tungsten collector. Comparison of these data with results obtained from the plane-parallel converters [2] shows good compatibility between them. At the same time the output power of TEC with plated collector made on the base of polycrystalline tungsten is at an average two times as big as the power of analogous TEC with polycrystalline niobium collector [2] within the range of high temperatures. So the use of tungsten collectors in high-temperature super-power thermo-emissive converters is far more advisible in case of durable maintainance of the high initial power. In this respect materials presented in [5] are of interest. There experimentally is determined the temperature conditions under which upon the tungsten surface in the oxygen atmosphere is formed the tungsten oxide with such order of electron bonds when oxygen can occupy only an upper part of its structure. Because of this upon the working surface of a tungsten electrode of the cesium TEC the oxygen monoatomic layer is formed and due to this exteremely positive results are obtained [4,5]. There was observed a very interesting effect of oxygen atoms migration towards the surface within the condenced layer formed on the oxidized tungsten surface by the mass transfer from the tungsten emitter. Oxygen is not soluble in tungsten, so despite the increase of a tungsten condensate thickness on the collector the oxygen monoatomic layer constantly remains in upper position, a strong shielding of a substrate takes place but its emissive properties are practically invariable.

Such polycrystalline tungsten collector of the cesium-plasma TEC with tungsten electrodes will function invariably practically endlessly and device resource will depend only on emitter layer's parameters [7].

As to tungsten (110) monocrystalline collector, it has a number of advantages:

- uniformity of emissive characteristics which causes the growth of the TEC's output voltage by 0,1 V;

- stability of emissive characteristics within the temperature range 900-1400 K while exists the cesium-plasma area between electrodes;

- absence of negative impact on the emitter on the part of the atmosphere between electrodes as well as the iodine cycle. But the monocrystalline (110) tungsten collector in contrast to niobium can not protect the emitter from the influence of active gases of device working volume and volt-amperic characteristics of such a TEC [8] are very sensible to admixtures condenced upon them and issued by the emitter and by the gap between electrodes. This process provokes the degradation of parameters of the TEC with monocrystalline tungsten collector. An analogous effect caused by gases produced during the nuclear fuel's exhalation is expected if these gases via emitter get to TEC working volume.

Therefore it is possible to say that in the thermo-emissive nuclear power plants the realization of monocrystalline tungsten collectors is a matter of future and it requires a higher level of technologies for electrogenerating elements and of vacuum hygiene.

Sukhumi Institute of Physics and Technology

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PHYSICS

Z.Kachlishvili, K.Jandieri

Hopf Bifurcation and the Possibility of Dynamic Chaos in Semiconductors under Impure Breakdown

Presented by Academician L.Buishvili, July 8, 1996

ABSTRACT. **On the bases of analysis of continuous oscillations appearing because of Hopf bifurcation in compensated semiconductor under impure electric breakdown condition it has been established that they appear in definite interval of basic impure concentration change under rather high compensations. Optic affect on the sample greatly expands the part of electric field changing which corresponds to continuous oscillations. There has been built the bifurcation diagram of the system on which the range of equilibrium dots of saddle-focus type where continuous as well as chaotic oscillations can appear is singled out. These last ones can have different character depending on whether random fluctuation of parameters or dynamic variables of the system or else the influence on them by some other periodic components will be taken into account or not.**

In our previous work [1] we have established necessary and sufficient condition of appearing persistent autooscillations in space-homogeneous compensated semiconductor under impurity electric breakdown.

$$\tau_D B_T^B N_D C > \frac{1+\alpha m}{\alpha} \tag{1}$$

where τ_D is delay time of electron temperature relatively to electric field change, B_T^B is coefficient of thermal recombination (the mark "B" everywhere and here indicates that corresponding meaning is taken in breakdown point). N_D is donors concentration and C is compensation degree; m is index of mobility dependence degree from electron temperature ($m = -1/2$, 0, 3/2 - under scattering of impulse on lattice oscillations on neutral and ionized atoms of impure correspondingly). $\alpha \equiv Z'_{oy}(y^*)y^*/Z^*$, where Z_0 is stationar meaning of undimensional electron temperature $Z=Te/T$ (T is lattice temperature), $y=E/E_B$ where E is electric field voltage. The mark "*" denotes equilibrium meanings of corresponding quantities.

The part of changing field in which the appearing of continuous as well as chaotic oscillations are expected is defined by the following unequations:

$$y^* - 1 < \frac{\alpha}{1+\alpha m}\cdot\frac{a_I^B}{\omega\mu_0 C z^{*m}} \tag{2}$$

where $a_I = A_I N_D$; A_I is collision (impact) ionization coefficient; μ_0 is mobility meaning under thermodynamic equilibrium. $\omega = 4\pi e N_D/\varepsilon$ where ε is dielectric penetration of the sample.

Thus the fulfilment of the condition depends on the character of hot electrons scattering and parameters entering in the left part of this equation. However it is worth noting that it is not so easy to set their optimal meanings as they are in complicated functional interrelations between each other. For example, the growth of C causes decreasing of B_T^B; the growth of N_D causes the decreasing τ_D and change of scattering character. Really it is evident from the formula [1] that the more the meaning of compensated degree the better is the condition of persistant oscillations appearance. But along with the increasing of C the breakdown meaning of electric field increases which decreases the meaning B_T^B . In spite of this if we take into consideration that the decreasing rate of B_T^B for high compensations is lower it can be suggested that the increasing of C promotes the fulfilling of condition (1). At the same time as it is seen from $\tau_D/\tau_e=f(Z_0)$ dependencies given in [1], (τ_e - is time of relaxation energy). The growth of compensation rate increases this relation. If we take into account that with increasing of C increases τ_e it can be finally proposed that the increasing of compensation rate increases τ_D and in this respect it also promotes the fulfillment of conditions for persistant oscillations. It is easy to make sure in breakdown range the role of impulse scattering increases by aucustic phonones. This is vividly seen under high compensations when the small change of C causes strong change of E_B [2]. Proceeding from this it can be concluded that the growth of compensation degree increases the role of impulse scattering on aucustic phonon, and this promotes the fulfillment of (1) and widens that part of field change which is defined by formula (2) and corresponds to persistent oscillations.

If we undertake analogous reasoning about the role of τ_D and N_D , finally it can be concluded that condition (1) will be fulfilled best of all in definite change interval N_D, under rather high compensations. This conclusion is in good agreement with experimental data obtained in [3,4]. For example, according to [3] persistant oscillations in GaAs are observed under concentration $N_D = 3\cdot10^{15}$ cm^{-3}, and under $N_D = 2\cdot10^{14}$ cm^{-3} such oscillations have no place. According to [4], the appearance of persistant oscillations in compensated semiconductors is expected in case of high compensations.

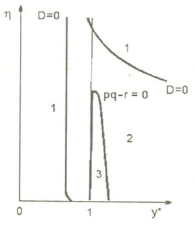

Fig. Quality picture of bifurcation diagram on the plane (y^*,η) in case of illuminate absence

On the base of corresponding selection of parameters of the system and analyses of discriminant of characteristic equation D, bifurcation diagram can be built on plane y^*,η, where $\eta = 4\pi L/\varepsilon SR$. Here L is the length of sample along the field, S is its cross section and R is load resistance switched in succession with sample. The quantitative picture of the diagram is given in Figure. It is illustrated that the range 1 corresponds to equillibrium dots of stable knots; range 2 corresponds to stable focuses, i.e. damped oscillations and the range 3 stands for saddle-focuses, i.e. continuous oscillations in semiconductor. The range 3 is defined with the help of Rauss-Hurvits condition: $pq-r<0$ (p,q and r are coefficients of characteristic equation). For our case they are always positive. The curves $pq-r = 0$ and $D = 0$ dividing different ranges define bifurcation meanings of parameter η for the given y^*. The bifurcation causing continuous autooscillations is of particular interest for us.

At the same time the formation of limited cycle from immovable point i.e.Hopf bifurcation takes place. The corresponding meaning η_{bif} is defined by the equation for η outgoing from the condition $pq - r = 0$ [1].

If we take into consideration that $\eta \sim L/\varepsilon SR$, bifurcation can be regarded by L, ε, S parameters of different samples in the condition of external parameter load resistance R constansy or which is more convenient, bifurcation on R for the given sample. The width of range 3 is defined by formula (2). It greatly depends on the meaning α . For instance, under $T = 4.2\ K$, $N_D = 10^{14}\ cm^{-3}$, $c = 0.9$ $\tau_D = 5\cdot 10^{-9}$ sec and under $\alpha = 1$ $y^* - 1 \approx 2.5\cdot 10^{-4}$, and under $\alpha = 1.9$ $y^* - 1 \approx 10^{-2}$ i.e. forty times more.

The formula (1) was obtained for the case of abscence of illuminating when the appearance of continuous oscillations is possible only after breakdown range of electric field. Because of optic influence on the sample the range 3 moves lefthand and considerably widen in direction of before breakdown meanings of electric field (i.e. in the direction of range where $y^* < 1$).

Correspondingly in this case continuous oscillations can appear both before and after breakdown. Optic affect can expand the range 3 hundreds and more times in direction of before breakdown meanings of electric field. As it was mentioned above the growth of α considerably expands indicated range aside to the after breakdown meanings of the field (i.e. aside to the range where $y^* > 1$). Both of these circumstances promote experimental observation of continuous autooscillations in semiconductors.

From the present studies it can be concluded that in conditions defined with the help of (1) and (2) chaotic oscillations can also appear in the system. Depending on whether random fluctuations of parameters of the system or dynamic variables or some periodic components affect on them will be taken into consideration, it can be possible to obtain various pictures of system transfer in chaotic regime.

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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PHYSICS

G.Jandieri, Zh.Diasamidze, V.Glonti

To the Theory of Particles Diffusion in a Turbulent Stream

Presented by Academician J.Lominadze, October 22, 1996

ABSTRACT. **This paper is devoted to the calculation of statistical state characteristics of scattered particles in a turbulent stream. The condition of probe particles is described by modifying Forkker-Planck equation and stochastic equation for a slowly moving spherical particle. Turbulent diffusion coefficient is calculated in the approximation of "free turbulence" in a pure spatial case for Gaussian temporal correlation function.**

The problem of velocity pulsations of randomly-inhomogeneous nonstationary medium and statistical description of turbulent diffusion has been given great attention lately. As in [1,2] general expression of effective permittivity tensor is obtained for arbitrary correlation tensor of velocity pulsation. The effect of negative turbulent diffusion is considered [3] in chaotically-inhomogeneous nonstationary media and the general expressions of longitudinal and transversal with respect to the stream motion, coefficients of turbulent diffusion are derived for arbitrary correlation function of velocity pulsation. Nonstationary problem of passive impurity concentration distribution is solved analytically [4] in a turbulent atmosphere flow. Numerical simulations of obtained results were carried out as applied to the ecology problems of environment pollution. Numerical data of turbulent diffusion coefficients are presented at different values of wind velocity and in the calm case at an altitude 100 m above the Earth surface.

This paper is devoted to the calculation of statistical characteristics of a state of scattered particles. The word "particle" means here an indelibly tagged fluid element, much smaller than the smallest length associated with the turbulence. The state of each particle is determined by its coordinate $\vec{r}$ and velocity $\vec{V}$. The consideration of the statistical characteristics of the track of a single indicator may present valuable information on the properties of the flow in the form of a distribution function $f(\vec{r},\vec{V},t)$. It is the probability of the selected particle, which at the initial moment has the coordinate $\vec{x}_0$ and the velocity $\vec{V}_0$, having after a certain period of time Δt the coordinate $\vec{x}$ and the velocity $\vec{V}$. Analytical calculations of the tensor coefficient of turbulent diffusion is carried out for a "free turbulence" (i.e. not taking into consideration the boundary influence and mechanism of energy supply).

We start from the stochastic equation for the slow motion of a spherical particle in a turbulent flow of an incompressible liquid: for the i-component of the particle velocity $\vec{V}(\vec{r},t)$ we have [5-7]

$$\frac{\partial V_i}{\partial t} = \alpha\left(\frac{\partial U_i}{\partial t} - \frac{2}{3}\nu\frac{\partial^2 U_i}{\partial x_i \partial x_j}\right) + b\left(U_i - V_i\right) + \frac{2}{3}\alpha\left(U_k - V_k\right)\frac{\partial U_i}{\partial x_k} +$$

$$+\left(\frac{\alpha b}{3\pi}\right)^{1/2}\int_{t_0}^{t} dt'(t-t')^{-1/2}\frac{d}{dt'}\left(U_i - V_i\right) \qquad (1)$$

where $U(\vec{r}(t),t)$ is the fluid velocity in those points where the particle exists (or, more exactly, following Tchen, in the neighborhood of the discrete particle), $\alpha = 3\rho_0/(2\rho + \rho_0)$, $b = 36\mu/R^2(2\rho + \rho_0)$; ρ_0 and ρ, are fluid and particle densities, respectively; μ and $\nu = \mu/\rho_0$ – are dynamic and kinematic viscosities, R is a particle radius. Further we will use the Lagrangian method for the description of the fluid motion. Furthermore, we have to consider that, in the other terms, $d/dt = \partial/\partial t + V_k\partial/\partial x_k$, since a coordinate system is moving with the discrete particle. The Equation (1) is drawing up similarly to the equation of Brownian motion. The term on the lefthand side is the particle acceleration, which is caused by turbulent pulsations. The latter are expressed through the derivatives of medium velocity with respect to coordinate and time. The last term is so-called Bassetian force, which takes account of the effect of the deviation in flow pattern from steady state. If the acceleration of particle occurs slowly, this term is negligible. Notice that in general case medium density is in the order of particle density. Therefore we will preserve all terms except the last one.

Let the condition of a probe particle in the turbulent flow be described by the modificatory Fokker-Planck equation:

$$\frac{\partial f}{\partial t} = -\frac{\partial}{\partial x_i}(A_i f) - \frac{\partial}{\partial V_i}(B_i f) + \frac{1}{2}\frac{\partial^2}{\partial V_i \partial V_j}(C_{ij} f) +$$

$$+\frac{1}{2}\frac{\partial^2}{\partial x_i \partial x_j}(D_{ij} f) + \frac{1}{2}\frac{\partial^2}{\partial x_i \partial V_j}(E_{ij} f) \qquad (2)$$

Probability distribution $f(\vec{r},\vec{V},t)$ changes in time and is a function of coordinate and the magnitude of velocity pulsations. Distribution function f might vary in form over different trails because it is a function of initial distributions which could differ. If we imagine that atmospheric pollution can be generated uniformly this case of variability would not occur.

In general case it is a random function of coordinate, velocity and time. The coefficients in Equation (6) are:

$$A_i = \lim_{\Delta t \to 0}\frac{1}{\Delta t}\int d\vec{x}d\vec{V}(x_i - x_{0i})f(\vec{x}_0,\vec{V}_0,t_0;\vec{x},\vec{V},t_0+\Delta t) = \lim_{\Delta t \to 0}\frac{<\Delta x_i>}{\Delta t} \qquad (3a)$$

$$B_i = \lim_{\Delta t \to 0}\frac{1}{\Delta t}\int d\vec{x}d\vec{V}(V_i - V_{0i})f(\vec{x}_0,\vec{V}_0,t_0;\vec{x},\vec{V},t_0+\Delta t) = \lim_{\Delta t \to 0}\frac{<\Delta V_i>}{\Delta t} \qquad (3b)$$

$$C_{ij} = \lim_{\Delta t \to 0} \frac{1}{\Delta t} \int d\vec{x} d\vec{V} (V_i - V_{0i})(V_j - V_{0j}) \cdot \\ \cdot f(\vec{x}_0, \vec{V}_0, t_0; \vec{x}, \vec{V}, t_0 + \Delta t) = \lim_{\Delta t \to 0} \frac{< \Delta V_i \Delta V_j >}{\Delta t} \tag{3c}$$

$$D_{ij} = \lim_{\Delta t \to 0} \frac{1}{\Delta t} \int d\vec{x} d\vec{V} (x_i - x_{0i})(x_j - x_{0j}) \cdot \\ \cdot f(\vec{x}_0, \vec{V}_0, t_0; \vec{x}, \vec{V}, t_0 + \Delta t) = \lim_{\Delta t \to 0} \frac{< \Delta x_i \Delta x_j >}{\Delta t} \tag{3d}$$

$$E_{ij} = \lim_{\Delta t \to 0} \frac{1}{\Delta t} \int d\vec{x} d\vec{V} (V_i - V_{0i})(x_j - x_{0j}) \cdot \\ \cdot f(\vec{x}_0, \vec{V}_0, t_0; \vec{x}, \vec{V}, t_0 + \Delta t) = \lim_{\Delta t \to 0} \frac{< \Delta V_i \Delta x_j >}{\Delta t} \tag{3e}$$

$< >$ denotes the averaging over all possible realization of the random parameters of a turbulent stream. Such statistical formulation of the turbulent diffusion coefficients (TDC) proceeds from Kolmogoroff equation describing random motion of the diffusible selected particle, which coordinate in time forms as a Markov process. Coordinate dispersion determines TDC. Utilizing Equation (1) let's find these coefficients. Particle velocity of nonstationary medium write as a sum of an average and fluctuating terms which are random function of coordinates and time:

$$\vec{V}(\vec{r}, t) = \vec{V}_0 + \vec{V}_1(\vec{r}, t), \quad U(\vec{r}, t) = U_0 + U_1(\vec{r}, t) \tag{4}$$

It is evident:

$$< \Delta x_i (\vec{r}, \Delta t) > = \int_0^{\Delta t} dt' < V_i(\vec{r}, t') > = V_{0i} \Delta t, \tag{5}$$

Hence,

$$A_i = V_{0i}, \tag{6}$$

i.e. particle mean velocity slightly changes during a small time interval Δt.

Inserting (4) into Equation (1), the linearized differential equation for a fluctuating component of the velocity of an impurity particle in wave space may be written as:

$$\frac{\partial V_{1i}}{\partial t} + \wp(k) V_{1i}(\vec{k}, t) = \alpha \frac{\partial U_{1i}}{\partial t} + Q(k) U_{1i}(\vec{k}, t) \tag{7}$$

The solution is

$$V_{1i}(\vec{k}, \Delta t) = V_{10i}(\vec{k}) e^{-\wp \Delta t} + e^{-\wp \Delta t} \int_0^{\Delta t} dt' e^{\wp t'} \cdot \\ \cdot \left[\alpha \frac{\partial U_{1i}(\vec{k}, t')}{\partial t'} + Q(k) U_{1i}(\vec{k}, t') \right] \tag{8}$$

where: $\wp(k) = b - i(\vec{k}\vec{V}_0)$, $Q(k) = b + \frac{2}{3} \nu \alpha k^2 - \frac{1}{3} i\alpha[(\vec{k}\vec{V}_0) + 2(\vec{k}\vec{V}_0)]$,

$V_{10i}(\vec{r}) = V_{1i}(\vec{r}, t = 0)$ is the particle velocity fluctuation at instant of time interval Δt. At small time interval Δt such as: $\wp\, \Delta t << 1$, expending the exponent into a series and keeping only the terms proportional to Δt, we obtain:

$$\Delta V_i(\vec{r}, \Delta t) \equiv V_i(\vec{r}, \Delta t) - V_{0i} = \Delta t[-bV_{10i}(\vec{r}) + (\vec{V}_0\nabla)V_{10i}(\vec{r})] +$$

$$+ \int_0^{\Delta t} dt' e^{bt'} \int_{-\infty}^{\infty} d\vec{k} \left[\alpha \frac{\partial U_{1i}(\vec{k}, t')}{\partial t'} + Q(k) U_{1i}(\vec{k}, t') \right] \qquad (9)$$

Particle velocity pulsations $\vec{V}_1$ are determined by the known statistical characteristics of nonstationary medium as it follows from Equations (8) and (9). Stochastic processes in the turbulent stream we suppose to be homogeneous and stationary with zero mean values $<U_1> = 0$. Hence, proceed from (9), coefficient $B_i = 0$.

Let's consider the tensor:

$$<\Delta V_i(\vec{r}_1, \Delta t)\Delta V_j(\vec{r}_2, \Delta t)> = 2b\Delta t \int_{-\infty}^{\infty} d\vec{k}'\, d\omega' W_{ij}(\vec{k}', \omega')\cdot$$

$$\cdot \frac{\alpha^2 \omega'^2 + Q^2(k')}{\left[\omega' + (\vec{k}'\vec{V}_0) + ib\right]\left[\omega' + (\vec{k}'\vec{V}_0) - ib\right]} e^{i\vec{k}'\vec{\rho}} \qquad (10)$$

$W_{ij}(\vec{\rho}, \tau)$ is a spatial-temporal correlation function of velocity field of the turbulent flow, $\vec{\rho} = \vec{r}_1 - \vec{r}_2$. This is a tensor of second rank. Integrating this expression over frequency and utilizing residue theory we obtain

$$C_{ij} = \frac{\alpha^2}{9} \int_{-\infty}^{\infty} d\vec{k} d\tau W_{ij}(\vec{k}, \tau) \left\{ 9\left[(\vec{k}\vec{V}_0)^2 - 2ib(\vec{k}\vec{V}_0) - b^2\right] + \right.$$

$$\left. + \frac{9}{\alpha^2}(b + \frac{2}{3}\nu\alpha k^2)^2 + \left[(\vec{k}\vec{V}_0) + 2(\vec{k}\vec{U}_0)\right]^2 \right\} e^{i\vec{k}\vec{\rho} - \left[b + i(\vec{k}\vec{V}_0)\right]\tau} \qquad (11)$$

This coefficient may be interpreted as the energy dissipation. Actually, on the one hand, coefficient of the turbulent diffusion is determined by the formula (3d), on the other hand utilizing similarity method, we have $<(\Delta x)^2> \sim C(\Delta t)^3$, $D \sim C(\Delta t)^2$, $<(\Delta V)^2> \sim C\Delta t$. Introducing characteristic spatial scale l for the probability distribution f (assuming, for example, $l^2 = <(\Delta x)^2>$) and eliminating time interval Δt, we obtain $<(\Delta V)^2> \sim (Cl)^{2/3}$, $D \sim C^{1/3} l^{4/3}$.

Let's consider the correlation function proposed by Taylor:

$$<\Delta x_i(\vec{r}_1, \Delta t)\Delta x_j(\vec{r}_2, \Delta t)> = \int_0^{\Delta t} dt' \int_0^{\Delta t} dt'' <V_i(\vec{r}_1, t')V_j(\vec{r}_2, t'')> =$$

$$= 2\int_0^{\Delta t} d\tau(\Delta t - \tau)\chi_{ij}(\vec{\rho}, \tau) \tag{12}$$

where $\tau = t' - t''$, $\chi_{ij}(\vec{\rho}, \tau)$. Correlation function of the particle velocity pulsation χ_{ij} express in terms of the tensor $W_{ij}(\vec{\rho}, \tau)$ .Proceed from the Equation (11) we have:

$$V_{1i}(\vec{k}, \omega) = \frac{-i\alpha\omega + Q(k)}{-i\omega + \wp(k)} U_{1i}(\vec{k}, \omega) \tag{13}$$

Consequently

$$\chi_{ij}(\vec{\rho}, \tau) \int_{-\infty}^{\infty} d\vec{k} d\omega \frac{\omega^2\alpha^2 + Q^2(k)}{\omega^2 + \wp^2(k)} W_{ij}(\vec{k}, \omega) e^{i\vec{k}\vec{\rho} - i\omega\tau} \tag{14}$$

Inserting equation (14) into (12), at $\Delta t >> \tau$ we get

$$D_{ij}(\vec{\rho}, \tau) = \int_0^{\infty} d\tau \int_{-\infty}^{\infty} dt' d\vec{k} \frac{Q^2(k) - \alpha^2\wp^2(k)}{\wp(k)} W_{ij}(\vec{k}, t') e^{i\vec{k}\vec{\rho} - \wp(k)|\tau - t'|} \tag{15}$$

It is easy to calculate the last coefficient of the Fokker-Planck equation using (5) and (9). If conditions $\tau << \Delta t << T$ are fulfilled and $b\Delta t << 1$ (T is a characteristic temporal scale of nonstationary medium velocity pulsations), neglecting the terms proportional to $\sim(\Delta t)^2$ we obtain:

$$E_{ij} = 2\left[-\alpha G_{ij}(\vec{\rho}, 0) + b\int_0^{\infty} d\tau G_{ij}(\vec{\rho}, \tau)\right] \tag{16}$$

The tensor

$$< U_{1i}(\vec{r}_1, t') V_{1j}(\vec{r}_2, t'') >= G_{ij}(\vec{\rho}, \tau) \tag{17}$$

is a new correlation function of the medium and particle velocities.

Let's connect tensor $G_{ij}(\vec{\rho}, \tau)$ with W_{ij}. For this purpose we will proceed from the expression (13):

$$G_{ij}(\vec{r}, t) = \int_{-\infty}^{\infty} d\vec{k} d\omega \frac{ia\omega + Q(k)}{-i\omega + \wp(k)} W_{ij}(\vec{k}, \omega) e^{i\vec{k}\vec{r} - i\omega t} \tag{18}$$

Integrating by frequency we get

$$G_{ij}(\vec{\rho}, \tau) = \int_{-\infty}^{\infty} d\vec{k} dt' [\alpha\wp(k) + Q(k)] W_{ij}(\vec{k}, t') e^{i\vec{k}\vec{\rho} - \wp(k)|\tau - t'|} \tag{19}$$

So all coefficients of the formula (3) are calculated.

Distribution function $f(\vec{r}, \vec{v}, t)$ is connected with particle concentration $N(\vec{r}, t)$ as:

$$N(\vec{r}, t) = \int d\vec{V} f(\vec{r}, \vec{v}, t) \tag{20}$$

and satisfies the following normalization condition:

$$\int d\vec{r} d\vec{V} f(\vec{r}, \vec{v}, t) = N_0$$

where N_0 is a total number of particles.

Taken into account (5), (11), (15) and (16), Equation (2) may be written in the form:

$$\frac{\partial f}{\partial t} = -A_i \frac{\partial f}{\partial x_i} + \frac{1}{2} C_{ij} \frac{\partial^2 f}{\partial V_i \partial V_j} + \frac{1}{2} D_{ij} \frac{\partial^2 f}{\partial x_i \partial x_j} + \frac{1}{2} E_{ij} \frac{\partial^2 f}{\partial x_i \partial V_j} \tag{21}$$

Integrate Equation (21) with respect to V in the approximation of "free turbulence" utilizing Equation (20) and Gauss theorem. The terms containing the derivatives with respect to velocity on the boundaries at infinity will be zero (in accordance with Tchen, turbulence has an infinite extent). Availability of the boundaries is preserved these terms). So we have

$$\frac{\partial N}{\partial t} + V_{0i} \frac{\partial N}{\partial x_i} = \frac{1}{2} D_{ij} \frac{\partial^2 N}{\partial x_i \partial x_j} \tag{22}$$

Knowledge of two-point Eulerian spatial-temporal correlation tensor of homogeneous stationary velocity fields

$$W_{ij}(\vec{\rho},\tau) = < U_i(\vec{r},t) U_j(\vec{r}+\rho, t'+\tau) >$$

allows us to calculate CTD; and solving the Equation (22) we can restore the mean distribution of impurity concentration.

In a pure spatial case and Gaussian temporal correlation function

$$\begin{aligned} W_{ij}(\vec{k},t) &= W_{ij}(t)\delta(\vec{k}) = W_{ij}(0)\exp\left(-\frac{t^2}{T^2}\right)\delta(\vec{k}) = \\ &= < U_1^2 > \exp\left(-\frac{t^2}{T^2}\right)\delta(\vec{k})\delta_{ij} \end{aligned} \tag{23}$$

proceed from the Equation (15) at $\alpha << 1$, $bT < 1$ for the isotropic diffusion coefficient we obtain

$$D = \sqrt{\pi} T < U_1^2 > \tag{24}$$

Georgian Technical University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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PHYSICS

A.Ugulava

Classification of Resonance Processes on Nonlinear Paramagnetic Systems

Presented by Corr. Member of Academy T.Sanadze, August 6, 1996

ABSTRACT. **NMR is studied in magnetics for low temperatures when a dynamic shift of the resonance frequency occurs. Nonlinear NMR resonance are classified and the limits of their realization are established.**

A dynamic shift of the precession frequency of nuclear magnetization at low temperatures in some magnetics [1,2] results in an essentially nonlinear character of NMR. A great deal of experimental and theoretical work has been devoted to the investigation of the magnetization motion under conditions of a dynamic frequency shift. However, taking into account the variety of motion types usually peculiar to nonlinearly oscillating systems one can suggest that the investigation of nonlinear NMR has not yet been completed.

Let us classify possible motion regimes in the given system. We shall use the classification of motion types in nonlinearly oscillating systems proposed by Volosov and Morgunov [3]. It is based on different time scales characterizing the system. According to this classification the following resonance motion regimes can be distinguished:

1) the system sticks in the resonance for a long time;
2) the system jumps through all possible resonance (i.e. it can be in the resonance for an infinitely short time);
3) the system turns from one state to another being in each of them for a certain finite time. These three types of the resonance motion regimes can be subjected to an approximate analytical description.

Let us assume that a periodic series of radio-frequency (RF) pulses is applied to the nuclear system

$$\vec{H}_1(t) = 2\vec{h}\, e^{i\omega t} \sum_{n=-\infty}^{\infty} f\left(\frac{t}{T} - n\right), \quad \vec{H}_1 \perp \vec{H} \tag{1}$$

$$\omega\tau >> 1, \ \tau << T$$

where $2h$ and ω are the amplitude and frequency of the variable field in the pulse peak, $f(t)$ is the pulse form function, τ and T are the pulse duration and pulse repetition period, respectively, $H = H_0 + H_s$ is the value of the constant magnetic field, H_s is the value of the hyperfine electron field at a nucleus. Let us direct the axis x along the total constant field and the axis x along the $\vec{H}_1$ field. Then the equation of motion for nuclear magnetization $\vec{m}$ in the coordinate system rotating about the axis z with the frequency ω is

$$
\begin{cases}
\dot{m}_x = \omega(m_z)m_y - \dfrac{m_x}{T_2} \\
\dot{m}_y = -\omega(m_z)m_x + \omega_1(t)m_z - \dfrac{m_y}{T_2} \\
\dot{m}_z = -\omega_l(t)m_y - \dfrac{m_z - m_0}{T_1}
\end{cases}
\tag{2}
$$

$$
\omega_l(t) = \eta\gamma H_l(t),\ \omega(m_z) = \omega_0 - \omega - \omega_p \frac{m_z}{m_0}
$$

where $\omega_0 = \gamma H$ is the unshifted NMR frequency, ω_p and m_0 are the equilibrium values of the dynamic frequency shift and magnetization, respectively, γ is the nuclear gyromagnetic ratio, η is the RF field amplification factor, T_1, T_2 are the longitudinal and transverse relaxation times.

To facilitate the investigation of equations (2) let us use the cylindrical coordinates according to the rule

$$
m_x = m_\perp \cos\varphi,\ m_y = m_\perp \sin\varphi,\ m_\perp > 0
$$

where $m_\perp$ is the transverse component of magnetization, φ is the rotation phase in the transverse plane. For convenience, let us take the pulse form to be ε-shaped. We obtain from (2)

$$
\begin{cases}
\dot{m}_z = -\omega_l(\lambda)m_\perp \sin\varphi - \dfrac{m_z - m_0}{T_l} \\
\dot{m}_\perp = \omega_l(\lambda)m_z \sin\varphi - \dfrac{m_\perp}{T_2} \\
\dot{\varphi} = -\omega(m_z) + \omega_l(\lambda)\dfrac{m_z}{m_\perp}\cos\varphi \\
\dot{\lambda} = \Omega
\end{cases}
\tag{3}
$$

where

$$
\omega_l(\lambda) = 2\pi\omega_l \sum_{n=-\infty}^{\infty} \delta(\lambda - 2\pi n) = 2\pi\omega_l \sum_{n=-\infty}^{\infty} \cos n\Omega t \tag{4}
$$

$$
\omega_l = \eta\gamma h \frac{\tau}{T},\ \Omega = \frac{2\pi}{T},\ \lambda = \Omega t.
$$

Let us regard the variable field and relaxation processes as a small perturbation, i.e. ω_l, $1/T_1$, $1/T_2 << 1/T$, ω_p. Then, according to the conventional terminology, M_{Hz} and $m_\perp$ can be considered as "slow" amplitudes and φ and λ as "fast" phases. System (3) describes a resonance interaction of two oscillating subsystems. In this case by the internal resonance in system (3) is meant fulfillment of the conditions

$$
\omega\left(m_z^{(n)}\right) = n\Omega \tag{5}
$$

where $m_z^{(n)}$ denotes resonance values of the longitudinal magnetization.

Let us introduce a parameter which is important in classification of nonlinearly oscillating processes - the time when the system is in resonance

$$\tau_2 = \left(\frac{d\omega}{dm_z} \dot{m}_z T_2 \right)^{-1} \tag{6}$$

It should be mentioned that $1/\tau_2$ is the frequency distance covered by a "creeping" resonance frequency in a time T_2.

Now let us classify resonance types of nonlinear NMR.

When the first resonance type - "sticking", is realized, the time τ_2 for which the system is in resonance is the longest time of the system – $\tau_2 >> T_2, T_1, 1/\omega_p, T$. Here Chirikov condition [4,5] of the absence of resonance overlapping is automatically fulfilled

$$K \approx \frac{d\omega}{dm_z} \dot{m}_z T^2 \approx \frac{T}{\tau_2} \frac{T}{T_2} << 1 \tag{7}$$

When (7) is satisfied, no dynamic stochasticity occurs in the system. In this case, as it has been mentioned in ref. [6], "slow amplitudes" m_z and $m_\perp$ and a slow phase $\varphi + \lambda n$ formed by two "fast" phases remain near their stationary values for a long time of the order of $2\pi/\omega_1$.

In the opposite limiting case $K >> 1$, $T << T_2$, T_1 the second motion type - resonance "jumping" is realized. This condition is actually fulfilled due to a high value of the "creeping" velocity of the resonance frequency i.e. at a sufficiently high nonlinearity in the system. Though in this case the resonance time of the system τ_2 is short, variation of the resonance frequency $\omega(m_z)$ even for such a short time is so large that the system can "jump" through the frequency distance of Ω order and fall into a resonance with another harmonic of the external field (4). If $K >> 1$, as it has been shown in ref. [7,8], during the "jump" further motion becomes unpredictable. This results in a dynamic stochasticity of motion.

Now let us proceed to the description of the third type of the resonance motion. Let us assume that T_2 and $\omega(m_z)$ (i.e. $T_2 << \tau_2, T, T_1, 2\pi/\omega_1$) are the shortest times in the system [9]

$$\omega_\perp^{(n)} = \omega_1 T_2 \sin\varphi^{(n)} m_z^{(n)} \tag{8}$$

$$\mathrm{ctg}\varphi^{(n)} = \omega(m_z^{(n)}) T_2$$

Substituting these expressions into the equation of the system (3), we obtain:

$$\dot{m}_z^{(n)} = -\omega_1^2 \frac{m_z T_2}{1 + \omega^2 (m_z^{(n)}) T_2^2} - \frac{m_z^{(n)} - m_0}{T_1} \tag{9}$$

which represents the Bloembergen-Parcell-Pound equation for a nonlinear case.

It should be mentioned that equation (9) describing nonlinear phenomena in NMR without any indication of the field of their application was used in a number of works [10,11]. And here we have shown that the application of this equation is valid only under condition of the sufficiently weak nonlinearity determined by the condition which, with allowance for (9), can take the form:

$$\omega_p << 2\pi / \omega_1^2 T_2^3 \tag{10}$$

We highly appriciate fruitful discussions with Acad. L.Buishvili who stimulated creation of this work.

Iv.Javakhishvili Tbilisi State University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

ASTRONOMY

T.Adeishvili, A.Gabeshia, I.Khwedelidze, I.Kopaliani

Shoemaker-Levi Comet Collision with Jupiter Planet and its Influence on the Earth Geomagnetic Field

Presented by Academician J.Lominadze, October 28, 1996

ABSTRACT. **The processes of Shoemaker-Levi comet and Jupiter collision have been observed. Hypothesis supposition is that these processes are caused by the flow of power particles (electrons and protons) from Jupiter radiation zones. Having reached the Earth magnetosphere they disturb its geomagnetic field.**

As it is known, the processes proceeding in a Solar system have influence on variations of some top layers parameters of the Earth atmosphere. On the other hand occurence of sky bodies in Solar system are not rare phenomena. Such alien bodies reach seldom the Earth as so-called external planets of Solar system play the role of protector. To confirm this there are craters on their surfaces, which are well seen in telescope. Sometimes some of them still can reach us which is proved by a fall of a known Tungus meteorit in the beginning of our century in Siberia.

It is known that in July 16, 1994 Shoemaker-Levi comet fell on a surface of Jupiter. It was fixed on the orbit of Jupiter two years ago. It has passed by Jupiter so close, that the comet's orbit has become dependent from a planet. Moreover the comet has passed a border of the fields of attraction of Jupiter, inside of which hightide-lowtide forces change the law of bodies' attraction. It was proved by supervision. 20 splinters of a comet the diameters of which were from hundred meters up to several km were fixed. On the 16th of July 1994 one of the splinters has come into the top layers of atmosphere of Jupiter with speed of 60 km/s. Others depending on their sizes and the weights, have soon reached the atmosphere of Jupiter. Penetration of a comet with such a speed and caused by it frictions form high temperature and radiation, during which according to some calculations energy 10^{25}-10^{28} erg is given off. For this reason it's possible to evaluate it as the phenomenon of a century. This collision has not directly affected the Earth; but the process was reflected on the top layers of atmosphere of the Earth, particularly, it influenced the processes in magnetosphere.

On the basis of an experimental material one of the possible gears causing disturbing is considered. Thus some assumptions are proposed. The record of the information took place in the observatory of our Institute which is located in Kumuri, Vani region, Georgia (x = 42.1, y = 42.3).

The observatory territory is optimally removed from large industrial centers and consequently the background of parasitic signals is minimum. Variations of the Earth magnetic field were registered by magnetogauge, which measured components of the field in three directions: east-west, north-south and vertical. Measurements were conducted continuously in two ranges: high-frequency 0.01-0.31 Hz and low-frequency 0.01-0.1 Hz. Sensitivity of the equipment in a high-frequency band was 0.01γ, and in low-frequency 0.1γ. The information on the magnetogauge is transmitted from special gauges, which represent magnetic solenoids and are equipped by special rods for amplification of intensity of a magnetic field voltage. The accepted signal is 8 times amplified. It is registered on a special register.

The investigated collision affected on our magnetosphere. Two hours after collision (02.00LT) geomagnetic pulsations of the type Pc3 were fixed. Moreover, remarkable increase of all geomagnetic field's components (= 40-50%) that is essential for night magnetosphere, was marked.

The occurence of Pc3 of geomagnetic pulsation at night magnetosphere is rather unexpected phenomenon. The pulsations of this type basically arise during the interaction of solar wind and magnetosphere.

At night the probability of their occurence is almost equal to a zero. It's necessary to note that pulsations of the similar form (first of all amplitude) up till now has not been fixed in our observatory. Geomagnetic pulsations of the type Pc3 were marked as a single pulse. It's essential, that magnetic pulsations of the type Pc3 were registered on east-western part of geomagnetic fields. Indignations of geomagnetic fields began to decrease rather quickly (40 min) and soon they came up to background parameters. Thus indignation was not casual. As it is known, magnetosphere of Jupiter has rather large size. Its magnetic field has an obviously dipole character near the planet (20 R_{ju}) and consists of radiating zones, in which highenergic electrons and protons, seized by a magnetic field are moving.

In the outermost areas magnetosphere is in average spread up to 60 R_{ju} and is deformed because of the rotation of the planet. In outermost areas there is an external magnetosphere (90-100), which is spead up to magnetopause where a magnetic field streamlined by solar wind charging particles defines a border between planets own magnetic field and interplanetary magnetic field. So the sizes of magnetopause change in due course [1]. Inside the magnetosphere of Jupiter in its radiating zones relativistic electrons and protons are seized. Simultaneously, in plasma the burst of particles in the plane of Jupiter's equator with energies up to 50 Mev for electrons and up to 70 Mev for protons are observed. Such particles leave magnetosphere and become additional sources of space beams in the solar system, both near and far from Jupiter. The flow of such particles is observed on the distances of 0.5-10 au from the Sun. The link of these processes with Jupiter is in modulation of an electronic flow with a period $\approx$10 hours, that coincides with a period of Jupiter's rotation [2]. We assume, that registered by us magnetic storm in July 15, which proceeded during a day and increased the intensity of geomagnetic fields in the average of 70-80 γ, was called by the movement of a comet in Jupiter's magnetosphere and pouring out additional flows of particles. Natural is that the movement of a comet in Jupiter's magnetosphere with supersound speed would call a sharp change of density of temperature and other parameters of plasma, that would in its turn cause without clash shock wave to appear. During its movement near the Jupiter's surface and at the moment of its fall a shock wave would appear. As far as we can not precisely define a time of generation of additional flow of pouring out particles, we consider, that they have arisen as a result of allocation of huge energy after collision [3,7]. Most of these particles would have had the speed, approximately equal to the light speed. Pouring out particles would have begun to move on Larmor orbit.

Let us evaluate the energy of the particles, which is necessary to achieve magnetosphere heights of the Earth for the time from collision till the registration of a signal.

The energy of relativistic protons the speed of which is $V = 2\cdot 10^5$ km/s, and equals to E=300 Mev, is 10^6 times more than the energy of protons of solar wind. It is possible to make the conclusion, that the disturbing which is registered by us in 2 hours after collision of a comet with Jupiter should call the arrival of additional flow of pouring particles. As to their speed and energy, which would be very important, here as the collision energy was $\approx 10^{28}$ erg could have called the affect noted above. We must also take into account that the flow of particles released was highenergical, which also proves the fact mentioned. Thus, the conducted experiment and analysis of the results showed the collision of a comet with a Jupiter (17.07.1994). The interaction was reflected in sharp changes of geomagnetic fields. Intensive magnetic storm was registered a few hours before the collision was observed which was probably connected with movement of a comet in the Jupiter's magnetosphere. After collision additional intensive flow of particles released and having reached the Earth magnetosphere excited magnetic pulsation of the type Pc3 (at hight hours). It called irregular disturbing in geomagnetic field of pulsation, that lasted for about an hour. Pulsation Pc3 was registered as a single pulse on West-East part (H_x) of magnetic field.

We consider that magnetic pulsation Pc3 was called by a flow of protons. Other variations of geomagnetic fields were called by electrons and heavy ions, first of all, oxygen ${}^{16}_{8}O^{+}$.

Kutaisi Technical University
Cosmophysical Scientific Research Institute

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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D.Akhobadze, M.Gverdtsiteli

Complex Molecules Algebraic Models Study by Modernized ANB-Matrices

Presented by Corr. Member of the Academy D. Ugrekhelidze, May 20, 1996

ABSTRACT. **Modelling of complex organic molecules and their study by modernized ANB - matrices method is considered. Diagonal elements of ANB - matrices represent atomic number of chemical elements, whereas nondiagonal ones - multiplicities of chemical bands.**

The age- long problem of organic chemistry - "structure - properties" can be solved by methods of correlation analysis, quantum chemistry and algebraic chemistry [1-3].

Let consider the molecule of lactame

$$\overline{CO - (CH_2)n - N} - (CH_2)m - R \tag{1}$$

where: R is variable substituent; n and m are integer numbers.

Contiguity matrices of molecular graphs and their various modifications are widely used in algebraic chemistry [4] . One type of such matrices are ANB - matrices; their diagonal elements represent atomic numbers of chemical elements, whereas nondiagonal elements - the multiplicity of chemical bonds [5]. For arbitrary XYV molecule ANB - matrix has a form

$$\begin{matrix} 1 & 2 & 3 \\ X & Y & V \end{matrix} \qquad \begin{Vmatrix} Z_X & \Delta_{XY} & \Delta_{XV} \\ \Delta_{XY} & Z_Y & \Delta_{YV} \\ \Delta_{XV} & \Delta_{YV} & Z_V \end{Vmatrix} \tag{2}$$

where: Z_X, Z_Y, Z_V are atomic numbers of X and Y and V chemical elements; Δ_{XY}, Δ_{XV} and Δ_{YV} represent multiplicities of chemical bonds between X and Y, X and V , Y and V.

It was found [4-6] that determinant value of ANB matrices can be considered as the topologic index for "structure - properties" correlation for some homologous series of organic compounds.

In case when in structure (1) n=6, m=2 and $R=C_6H_5$, the range of ANB - matrix is equal to 38; naturally the determinant has the same range, so calculations would be very labour - consuming.

We have elaborated a simple model, which reproduces the specific character of (1) system and at the same time is less labour - consuming.

Consider, that n=1 and m=1; R is variable substituent. In this case (1) system takes a form:

$$\overline{CO - CH_2 - N} - CH_2 - R \tag{3}$$

Conventionally this system can be written as:

$$B - L - R \tag{4}$$

where: B corresponds to fragment $\overline{CO-CH_2-N}-$; L to - CH_2; R - to variable substituent.

Let denote

$$Z_B = \Sigma Z_b \qquad (5)$$

$$Z_L = \Sigma Z_l \qquad (6)$$

$$Z_R = \Sigma Z_r \qquad (7)$$

where Z_b, Z_l and Z_r correspond to the atomic numbers of elements in fragments B, L, and R. For system (4) the modernized ANB-matrix ($A\tilde{N}B$) can be written as:

$$\begin{Vmatrix} Z_B & 1 & 0 \\ 1 & Z_L & 1 \\ 0 & 1 & Z_R \end{Vmatrix} \qquad (8)$$

and its determinant can be calculated by formula:

$$\Delta_{ANB} = Z_B(Z_L \cdot Z_R - 1) - Z_R \qquad (9)$$

If we take into account, that in case (3), Z_L=8, (Z_C+Z_H) and Z_B=29, ($2Z_C$+Z_0+$2Z_H$+Z_N), formula (9) will take a form:

$$\Delta_{A\tilde{N}B} = 57 Z_R - 29 \qquad (10)$$

By this and analogous formulae we can calculate topological indices - the values of $\Delta_{A\tilde{N}B}$ and investigate the correlation $W \approx \Delta_{A\tilde{N}B}$ (W represents some physical, chemical, biologic and other properties of the system).

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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ORGANIC CHEMISTRY

I.Fiolia, L.Asatiani, M.Pochkhidze

The Regio and Stereodirection of n-Butanethiol's Nucleophilic Addition to Ferrocenylacetylene

Presented by Corr. Member of the Academy G.Chivadze, July 6, 1995)

ABSTRACT. **The reaction of nucleophilic addition of n-butanethiols to Ferrocenylacetylene in DMFA medium has been studied.**

It has been established that reaction goes nonregiospecifically but stereospecifically. The product of α- and β-additions has been obtained. However, the product of β-additions has been mainly only with cis-configuration obtained (≈75%).

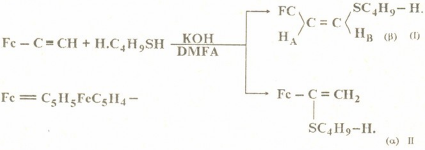
The structure and composition of received compounds have been established by IR-, and PMR-spectroscopy and elemental analysis.

Earlier [1] we have investigated the stereochemistry of tertiary-C_4H_9SH nucleophilic addition to ethynilferrocene [2]. It was established, that the reaction goes regiospecifically, with exclusive formation of β-addition products, but not stereospecifically, with reception of cis and trans isomers (60:40). For finding out how the replacement of tertiary-C_4H_9 by normal-C_4H_9 will affect the stereochemistry of reaction in the present work we have investigated normal-C_4H_9SH's interaction with ferrocenylacetylene.

Thiolation of ferrocenylacetylene [2] was carried out in dimethylformamide (DMFA) and caustic potash under the argon's atmosphere.

The course of reaction was watched with the help of thin-layer chromatography. For the end of reaction long heating ≈35 hours is required at 80°C.

PMR-spectroscopic analysis of products, allocated from a reactionary mix has shown, that the result of reaction are two compounds I and II, i.e. the addition of normal-C_4H_9SH, as against tertiary-C_4H_9SH, goes not regiospecifically, but stereospecifically. The addition occurs both at α- and β-sites, however mainly compound (I) (≈75%) with a cis-structure is formed.



In the PMR spectrum of a received product, after purification by preparatory thin-layer chromatography in hexane, there are two doublets of vinyl protons with the chemical shifts for H_A and H_B protons: δ_{H_A} 6.15 p.p.m., δ_{H_B} 6.00 p.p.m. with the 10.7 Hz spin-spinic interaction constant [3], and the singlet of last vinyl protons with chemical shifts $=CH_2$ 6.27 p.p.m. These data confirms the structure of compounds (I) and (II) respectively. The 1:3 ratio of integral intensities of the $=CH_2$ and CH=CH protons shows, that the compound (I) (≈75%) will be mainly formed.

In the infra-red spectrum of a mix of compounds (I, II) frequencies in 2150-2170 cm^{-1} (appropriate to C≡C bond oscillations) and around 3272 cm^{-1} (appropriate to ≡CH bond oscillations) are not available, but there are some lines around 1596 cm^{-1}, corresponding to oscillation of C=C bonds in compound (I). Also we have some characteristic frequencies near 3096 cm^{-1}, 1412 cm^{-1}, 880 cm^{-1}, appropriate to oscillations of $=CH_2$ bonds and the frequencies around 1640 cm^{-1}, appropriate C=C bond oscillations, that confirms the structure of the compound (II).

Such direction of the reaction, i.e. addition both in α- and β-sites, in difference from the tertiary buthylmercaptane's (I) addition, which because of spatial obstacles, adds only in β site, corresponds to the distribution of electronic density in the molecules of Ferrocenylacetylene, which favours the α- addition.

Our results are coordinated with data being available in the literature about connection of unbranched aliphatic thiols (including n-butanethiole) to the phenyl and p-tolilacetylene [4,5]. A little bit greater percent (≈25%) of α-isomer (II), in our case may be explained by considerably large positive inductive effect of the Ferrocenyl group in compare with the phenyl [6], which promotes the α-addition in a greater degree.

It is necessary to note also, that β-adducts (I) have a solely cis-structure, that corresponds to data that the reaction of I thiolation of alkines always goes by the rule of antiaddition irrespective from dissolvent's nature [5].

Addition of n-butanethiol to ferrocenylacetylene-2

Caustic potash (KOH) (0.0001M) and 4ml methanol were added to the solution of ferrocenylacetylene (0.0012 M) in 10 ml of dimethylformamide n-Butanethiol (0.0012 M) in 1 ml absolute DMFA was added through the argon current. After addition of whole quantity of thiol the control by the thin-layer chromatography in hexane has shown that the reaction hasn't begun. Temperature was increased in the reaction area. We heated up 10 h. at 40-50°C and 15 h. at 60-70°C, then we added a few drops hexamethylphosphortriamide at the same temperature and then continued heating at 80°C till possible reduction of quantity of initial substance (≈35h). We have added the diethylether to a solution, then a little water. A water layer was several times extracted by diethylether, have washed out by water, dried with $MgSO_4$. After removal of dissolvent in vacuum the product have dried up in vacuum exicator.

Thus we received dark-brown mass, crystallized at cooling, which after clearing by preparatory thin-layer chromatography on the non-fixed layer of Al_2O_3 in hexane by data of PMR spectra contains ≈75% β-isomer (I) with cis-structure and ≈25% α-isomer (II). An output was 60%.

Found %: C 64.32; H 6.50; S 11.13. $C_{16}H_{20}FeS$.

Calculated %: C 64.00; H 6.67; Fe 18.67; S 10.67

The PMR spectra were written by the apparatus Bruker WM-250 with working frequency 250 MHz in a solution $CDCl_3$.

The infra-red spectra were written by the apparatus Specord M-80 in a thin layer.

Tbilisi I.Javakhishvili State University.

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ORGANIC CHEMISTRY

M.Gverdtsiteli, N.Kobakhidze

Algebraic Investigation of Cycloalkanes

Presented by Corr. Member of the Academy D.Ugrekhelidze, 14 January, 1995

ABSTRACT. **Modifications of contiguity matrices of molecular graphs - ANB- and EP- matrices were used for algebraic investigation of cycloalkanes. In terms of this method "structure-property" correlation was found.**

Contiguity matrices of molecular graphs and their various modifications are efficiently used for algebraic characterization of organic molecules and their transformations [1]. ABN- and EP- belong to such type of matrices.

The diagonal elements of ABN- matrices represent atomic numbers of chemical elements, whereas nondiagonal ones- the multiplicities of chemical bonds [2]. For arbitrary ABC molecule ABN- matrix has a form:

$$\begin{matrix} 1 & 2 & 3 \\ A & B & C \end{matrix} \qquad \left\| \begin{matrix} Z_A & \Delta_{AB} & \Delta_{AC} \\ \Delta_{AB} & Z_B & \Delta_{BC} \\ \Delta_{AC} & \Delta_{BC} & Z_C \end{matrix} \right\| \qquad (1)$$

where: Z_A, Z_B, Z_C are atomic numbers of A, B and C chemical elements; Δ_{AB}, Δ_{AC} and Δ_{BC} represent multiplicity of chemical bonds between A and B, A and C, B and C.

Table

The values of determinants of ABN- and EP-matrices (Δ_{ANB} and Δ_{EP}) and corresponding standard entropies (S^0_{298})

Cycloalkane	S^0_{298}	Δ_{ANB}	Δ_{EP}
Cyclopentane	70.00	726	$1.2 \cdot 10^5$
Methcyclopentane	81.24	2596	$1.3 \cdot 10^6$
Ethycyclopentane	90.42	9658	$1.3 \cdot 10^7$
Propycyclopentane	99.73	36036	$1.3 \cdot 10^8$
Cyclohexane	71.28	2700	$1.3 \cdot 10^6$
Methylcyclohexane	82.06	9660	$1.3 \cdot 10^7$
Ethylcyclohexane	91.44	35940	$1.4 \cdot 10^8$
Propylcyclohexane	100.27	134100	$1.4 \cdot 10^9$

The diagonal elements of EP- matrices represent electronegativity of chemical elements, whereas nondiagonal ones- the polarity of chemical bonds[3]. For arbitrary ABC molecules EP- matrix has a form:

$$\begin{matrix} 1 & 2 & 3 \\ A & B & C \end{matrix} \qquad \left\| \begin{matrix} \chi_A & \mu_{AB} & \mu_{AC} \\ \mu_{AB} & \chi_B & \mu_{BC} \\ \mu_{AC} & \mu_{BC} & \chi_C \end{matrix} \right\| \qquad (2)$$

where: χ_A, χ_B and χ_C are electronegativities of A, B and C chemical elements; μ_{AB}, μ_{AC} and μ_{BC} represent polarities of A – B, A – C and B – C chemical bonds.

The values of standard entropies (S^0_{298}) [4] and values of determinants of corresponding ABN- and EP- matrices are given in Table.

The plot of $S^0_{298} \sim \lg\Delta_{ANB}$ and $S^0_{298} \sim \lg\Delta_{EP}$ dependence are constructed on computer:

$$S^0_{298} = 14.546 \ \lg\Delta_{ANB} + 27.808 \quad (3)$$

$$S^0_{298} = 8.266 \ \lg\Delta_{EP} + 27.118 \quad (4)$$

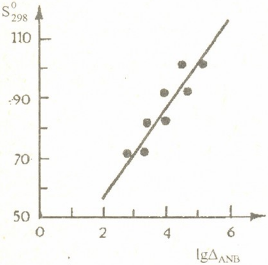

Fig. 1. The plot of dependence $S^0_{298} = 14.546 \ \lg\Delta_{ANB} + 27.808$ for some cycloalkanes.

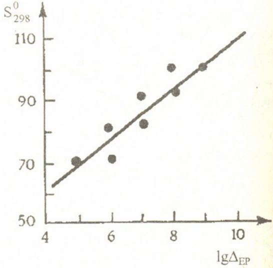

Fig. 2. The plot of dependence $S^0_{298} = 8.266 \ \lg\Delta_{EP} + 27.118$ for some cycloalkanes.

These plots are given in Figures 1 and 2. The correlation coefficient r in both cases is equal to r=0.997, so excellent correlations are observed.

Thus $\lg\Delta_{ABN}$ and $\lg\Delta_{EP}$ can be considered as topologic indices for "structure-property" correlation in homologous series of cycloalkanes.

I. Javakhishvili Tbilisi State University

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ORGANIC CHEMISTRY

E.Gelashvili, N.Iashvili, S.Kiladze

Synthesis of Some Bromosubstituted Acetylenic Epoxides

Presented by Corr. Member of the Academy D.Ugrekhelidze, June 21, 1996

ABSTRACT. **On the base of acetylenic carbinoles some bromosubstituted acetylenic epoxides were obtained. The structure of these compounds were determined by physico–chemical methods.**

Due to the oxide ring high reactivity, epoxides attract attention of researchers. They are widely used in different fields of the national economy.

Epoxides containing unsaturated bonds besides the oxide ring, are especially important. Including two reaction centres such compounds are widely used to synthesize organic compounds of various types.

In order to study the halogenation reaction we undertook bromination of some $HC \equiv C - CRR'O - CH2 - \underset{\diagdown O \diagup}{CH_2 - CH_2}$ type acetylenic oxyranes. At first dibromdioxane was used as a halogenetic agent. The reaction was carried out at 65–70^{0}C. It turned out that the substitution of the acetylenic hydrogen by bromine took place but only 3–5% yield from theoretical. Under the harder conditions the whole conversion of the reaction mixture into pitch was occured, whereas at low temperature the course of the reaction was practically prevented. Therefore we changed halogenating agent to potassium hypobromite [1].

Yield of bromosubstituted oxyrane in this case was proved to be comperatively better (20–25%):

$$HC \equiv C - CRR'O - CH_2 - \underset{\diagdown O \diagup}{CH_2 - CH_2} \xrightarrow{KOBr} BrC \equiv C - CRR'O - CH_2 - \underset{\diagdown O \diagup}{CH_2 - CH_2}$$

R=R'=H (I); R=i–C_3H_7, R'=H (II); R=R'=CH_3 (III):
R=CH_3, R'=C_2H_5 (IV); R=R'= –$(CH_2)_5$ – (V)

IR spectra of the synthesized compounds show $C \equiv C$ and $\underset{\diagdown O \diagup}{C - C}$ bands at 2220–2240 and 3030–3050 cm^{-1} respectively.

Due to the double excess of hypobromate along with the major products the following compounds were separated:

$CHBr=CBr – CRR'O – CH_2 – CH_2 – COOH$
R=R' =H (VI); R=i–C_3H_7, R' =H (VII); RR' = –$(CH_2)_5$ (VIII):

Formation of these products may be explained by the oxide ring isomerization into ketone and then its oxidation into the corresponding acid under the reaction.

IR spectra of the aforesaid compounds differing from other synthesized compounds, vibration is seen C=CV band at 1600–1620, C=0 band at 1750–1760 and OH band at 3400–3600 cm^{-1}.

To define the structure of the obtained bromosubstituted epoxides the approaching synthesis was carried out according to the scheme:

Table

N	Yield %	Boild. temp. $^0Cp_{mm}$	d_4^{20}	n_D^{20}	MR_D Found	MR_D Calcu–latedt	Found, % C	Found, % H	Found, % Hal	Formula	Calculated, % C	Calculated, % H	Calculated, % Hal
I.	46.3	80/2	1.4574	1.4847	37.5	37.4	38.2	3.9	40.6	$C_6H_7O_8Br$	37.4	3.7	41.2
II.	32.1	106–109/6	1.2820	1.4421	48.0	48.7	45.9	5.0	35.8	$C_9H_{13}O_2Br$	46.3	5.6	34.3
III.	39.2	110–112/22	1.3460	1.4843	46.5	46.7	44.3	4.7	35.2	$C_8H_{11}O_2Br$	43.9	5.0	35.8
IV.	43.3	122/20	1.2850	1.4428	51.3	51.5	45.6	6.2	34.9	$C_9H_{13}O_2Br$	46.1	6.0	34.2
V.	31.8	94/6	1.2500	1.4760	58.5	58.1	51.3	5.0	31.6	$C_{11}H_{15}O_2Br$	50.9	5.8	30.9
VI.	29.6	76–77/2	1.6293	1.4465	47.2	48.3	24.7	3.0	55.9	$C_6H_8O_3Br_2$	25.0	2.8	55.5
VII.	21.7	111–113/1	1.3381	1.4940	54.1	53.3	43.1	5.8	32.9	$C_9H_{13}O_3Br_2$	43.4	5.2	32.1
VIII.	5.0	158/2	1.6412	1.5490	69.0	69.3	37.3	4.6	44.0	$C_{11}H_{16}O_3Br_2$	37.1	4.4	44.9
IX.	50.0	52/3	1.3598	1.4879	54.0	53.3	37.0	5.3	44.6	$C_9H_{14}O_2BrCl$	37.6	4.6	45.2
X.	45.6	58/3	1.3645	1.4903	57.1	57.9	40.5	4.9	41.5	$C_9H_{14}O_2BrCl$	40.0	5.2	42.8

$$HC\equiv C-CRR'OH \xrightarrow{KOBr} BrC\equiv C-CRR'OH \xrightarrow{ClCH_2-\overset{O}{CH-CH_2}}$$

$$\longrightarrow BrC\equiv C-CRR'OCH_2-\underset{OH}{CH}-CH_2Cl$$

$$\xrightarrow{KOH} BrC\equiv C-CRR'OCH_2-\underset{O}{CH-CH_2}$$

(IX, X) (III, IV)

R=R' =CH_3 (III, IX); R=CH_3, R' =C_2H_5 (IV, X)

IR spectra were recorded on a UR 20 spectrometer with NaCl, LiF and KBr prisms. Starting acetylenic epoxides were obtained according to the known procedure [2,3].

1,2–Epi–4–oxa–7–bromoheptin–7 (1). 4 ml of bromine was dropwise added to a solution of 18 g of granular potassium alkali in 25 ml of water at 35 –40^{0}C and was mechanically stirred. Then potassium hypobromite solution 7g (0,7 M) of acetylenic oxyrane –1,2–Epi–4–oxaheptin–1 in ether was added at –15–10^{0}C, stirred for 2–2.5 h, extracted with ether and dried over Na_2SO_4. Frequent distillation of the reaction product gave colourless oily compound (I), b.p. 80^0/2mm Hg. Yeld 6g.

Compounds (II–V) were obtained in the similar manner.

Compound (VI) was synthesized under conditions similar for (I), only with the double–excess of bromine. From 4 ml of Br_2 and 3.5 g (0.35 M) of 1.2–Epi–4–oxaheptin–1 were prepared two compounds (I) b.p. 80–81^0/2 mm Hg, d_4^{20} 1.4570; n_D^{20} 1.4852. Yield 2.8 g.

In analogy with (VI) compounds (VII) and (VIII) were obtained.

1–Bromo–3,3–dimethyl–4–oxa–7–chloro–6–hydroxyheptin–1 (IX). 11,1 g of epichlorohydrin was dropwise added to 10 g of bromo-substituted dimethylacetylenylcarbinol (b.p. 58–60^0/10 mm Hg) under ice cooling with the catalytic amount of boron trifluoride etherate, then stirred for 3.5 h at 60–65^0 C. Frequent distillation of the reaction mixture gave compound (IX), b.p. 52^0 /mm Hg. Yeld 7.8 g.

From the compounds (IX) and (X) with the approaching synthesis (III) and (IV) were obtained according to the known procedure [2,3].

I. Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

ORGANIC CHEMISTRY

L.Baramidze, M.Gverdtsiteli

Algebraic Characterization of the Reactions of Aminosilanes with Tertiary Acetylene Carbinols

Presented by Corr. Member of the Academy L.Khananashvili, 29 May, 1996

ABSTRACT. **The reactions of aminosilanes with tertiary acetylene carbonols are characterized in terms of ANB-matrix method; diagonal elements of ANB matrices represent atomic numbers of chemical elements, whereas nondiagonal ones-multiplicities of chemical bonds.**

The reactions of tertiary acetylene carbonols with aminosilanes are one of the interesting processes in acetylene chemistry [1]. The concrete reaction is brought below:

$$
\begin{array}{c} CH_3 \\ | \\ HC \equiv C-C-OH \\ | \\ C_2H_5 \end{array}
\xrightarrow[- NH(C_2H_5)_2]{+(CH_3)_2SiH[N(C_2H_5)_2]}
\begin{array}{c} CH_3 \quad\quad CH_3 \\ | \quad\quad\quad | \\ HC\equiv C-C-O-Si-H \\ | \quad\quad\quad | \\ C_2H_5 \quad\quad CH_3 \end{array}
\qquad (1)
$$

Contiguity matrices of molecular graphs and their various modifications are efficiently used for algebraic characterization of molecules and their transformations in organic chemistry [2,3]. One type of such matrix is ANB - matrix. Its diagonal elements represent atomic numbers of chemical elements, whereas nondiagonal ones - the multiplicities of chemical bonds [4]. For arbitrary XYV molecule ANB - matrix has a form:

$$
\begin{array}{c} \\ XYV \\ \\ \end{array}
\quad
\begin{array}{c} 1\ 2\ 3 \\ \left\| \begin{array}{ccc} Z_X & \Delta_{XY} & \Delta_{XV} \\ \Delta_{XY} & Z_Y & \Delta_{YV} \\ \Delta_{XV} & \Delta_{YV} & Z_V \end{array} \right\| \end{array}
\qquad (2)
$$

where: Z_X , Z_Y , Z_V are atomic numbers of X, Y, V -chemical elements; Δ_{XY}, Δ_{XV}, Δ_{YV} represent multiplicities of chemical bonds between X and Y, X and V, Y and V.

It must be mentioned, that for large molecules calculations are very labour-consuming. We have elaborated a simple modelreproducing specific character of systems and at the same time it is less labour-consuming. For example, acetylene carbinol can be written as:

$$
\begin{array}{c} A^2 \\ | \\ B^3 - C^1 - D^5 \\ | \\ E^4 \end{array}
\qquad (3)
$$

where: A represents CH_3 , B - HC C, E - C_2H_5 and D - OH.

Corresponding modernized ANB-matrix (ANB) has a form:

$$\begin{Vmatrix} 6 & 1 & 1 & 1 & 1 \\ 1 & Z_A & 0 & 0 & 0 \\ 1 & 0 & Z_B & 0 & 0 \\ 1 & 0 & 0 & Z_E & 0 \\ 1 & 0 & 0 & 0 & Z_D \end{Vmatrix} \tag{4}$$

where: $Z_A = \Sigma Z_a$, $Z_B = \Sigma Z_b$, $Z_E = \Sigma Z_e$, $Z_D = \Sigma Z_d$ are the sums of atomic numbers of chemical elements with A, B, E and D fragments. Aminosilanes can be written analogously.

The matrix notations of (1) reaction has a form:

$$\begin{Vmatrix} 6 & 1 & 1 & 1 & 1 & 0 & 0 & 0 & 0 & 0 \\ 1 & 9 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 1 & 0 & 13 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 1 & 0 & 0 & 17 & 0 & 0 & 0 & 0 & 0 & 0 \\ 1 & 0 & 0 & 0 & 9 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 14 & 1 & 1 & 1 & 1 \\ 0 & 0 & 0 & 0 & 0 & 1 & 9 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 1 & 0 & 9 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 1 & 0 & 0 & 1 & 0 \\ 0 & 0 & 0 & 0 & 0 & 1 & 0 & 0 & 0 & 41 \end{Vmatrix} \rightarrow \begin{Vmatrix} 6 & 1 & 1 & 1 & 1 & 0 & 0 & 0 & 0 & 0 \\ 1 & 9 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 1 & 0 & 13 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 1 & 0 & 0 & 17 & 0 & 0 & 0 & 0 & 0 & 0 \\ 1 & 0 & 0 & 0 & 8 & 1 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 1 & 14 & 1 & 1 & 1 & 0 \\ 0 & 0 & 0 & 0 & 0 & 1 & 9 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 1 & 0 & 1 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 1 & 0 & 0 & 9 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 42 \end{Vmatrix} \tag{5}$$

Calculations show, that the value of determinant of ANB - matrices, describing (1) reaction increased. It is the algebraic criterion (in terms of ANB - matrix method) for this process.

I. Javakhishvili Tbilisi State University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

PHYSICAL CHEMISTRY

M.Gabunia, N.Kedrina, A.Berlin

Demonstration of Ideal Mixing Conditions in the Continuous Reactor by Realization of Formaldehyde's Polymerization

Presented by Corr. Member of the Academy L.Khananashvili, June 19, 1996

ABSTRACT. **The indicator method showed, that in the continuous reactor at the polymerization (and without it) of gaseous formaldehyde the reactor response corresponds to the regime of ideal mixing in the wide range of concentration of polyformaldehyde. In the various series of trials the monomer (when polymerization did not proceed) and polymer (at the polymerization performance) served as indicators.**

The process reflecting chemical reaction may have periodic, halfperiodic or continuous character. The continuous processes in comparison with others possess several advantages causing increased interest of scientists. Nevertheless, the list of scientific publications, devoted to the investigation chemical processes proceeding in continuous systems, is poor, especially, in the sphere of polymerization, kinetics and catalysis. Exploration of such systems, where the continuous delivery of initial compounds and at the same time disloading of reaction products are studied, may be carried out in two principally different containers: in the container of ideal elimination and the container of ideal mixing.

In the present work the polymerization of formaldehyde (FA) is discussed in the container of ideal mixing i.e. in the continuous reactor of ideal mixing (CRIM). The interest in this process is also stipulated by wish to meditate more deeper its kinetic and technological scheme of synthesis of polymer-polyformaldehyde (PFA), which has an industrial significance.

The polymerization was carried out in CRIM at the temperature 30°C. The principal scheme of polymerization installation resembles the scheme described in publication [1]. The volume V of the reactor liquid phase was equal to 0.064 L. The gaseous FA, purified from contaminants and obtained via thermal decay 175-180°C of α-polyoxymethylene, was put into the reactor with constant rate. The delivery of solvent (toluol) was realized together with catalyst $(C_4H_9)_3SnOCOCH_3$-tin tributilacetate (TTA) by the rate u=0.016 L/min, but concentration of catalyst made up $8\cdot10^{-6}$ M/L. The residence time of polymerization mass in CRIM was τ=4 min. The conditions have been selected so, that the delivery of the solvent and elimination of polymerization mass from CRIM was proceeded with the same rate. Toluol was purified and dried with standard methods [2]. TTA with qualification "analytically pure" was characterized by indices corresponding to the reference data. The solution of catalyst in toluol was prepared in the dry argon's medium. The content of FA in the reaction mixture was determined by sulfite method [3]. The washing off of the obtained polymer was accomplished with methanol and ether, then it was dried in vacuum at the room temperature to the constant weight.

In order, that the data obtained due to investigation result of the mechanism and polymerization kinetic of gaseous FA in CRIM may be considered confidential, it was primarily necessary to establish dynamic characterization of reactor i. e. whether reactor was working in the regime of ideal mixing. In the real reactors deviation from this regime often takes place, so that in our case proceeding of process is complicated by its heterogenic character. In order to establish process character, the indicator method was used [4,5]. As indicator monomer (when polymerization did not proceed) and polymer (during polymerization) were used.

In the case of ideal mixing for the indicator component the distributive function versus time has an exponential character:

$$X = X_0 e^{-\frac{u}{V}t} \qquad (1)$$

where X_o and X - relative are initial and current concentrations of indicator; u - reflects the rate of solvent delivery; V - means the volume of liquid phase in the reactor; t - is the time.

From the equation (1) it follows that

$$\ln\frac{X}{X_o} = -\beta t \qquad (2)$$

where $\beta = \frac{u}{V}$ =and is reciprocal dimension of τ - residence time of polymerization mass in the reactor i.e. $\tau = \frac{V}{u}$ and $\beta = -\frac{1}{\tau}$. In our case $\beta = \frac{1}{4} = 0.25\ \text{min}^{-1}$.

The mixing character was primarily established in the system, where polymerization did not proceed. For this purpose monomer FA was used as indicator. In this series of trials monomer was delivered to the solvent (without catalyst) until the stationary state (which is characteristic for reactor of such type) would be installed in the system. Then the delivery of monomer would be stopped, the probe was taken out and concentration of FA was determined. The model of relationship obtained as follows is given in Fig. 1 (curve 1).

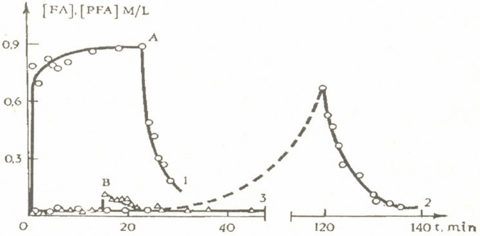

Fig. 1. Variation of FA (1) and PFA (2,3) concentration dependence on time
A - stopping the delivery of monomer,
B - addition of polymer,
--- accumulation of polymer.

In case of using as indicator the monomer equation (2) is transferred, so:

$$\ln\frac{[FA]}{[FA]_0} = -\beta\, t \qquad (3)$$

where $[FA]_0$, and $[FA]$ are initial and current concentrations of monomer, respectively.

As it is shown from Fig. 2. (line 1), $\ln\frac{[FA]}{[FA]_0} - t$ relationship is linear, which on the basis of the equation (3) indicates, that the reactor is working in the regime of ideal mixing. It should be noted, that calculated from this relationship the value of β is equal to 0.29 min^{-1}, which is in good correspondence with real value 0.25 min^{-1}.

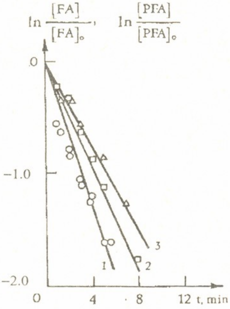

Fig. 2. Dependence of $\ln\frac{[FA]}{[FA]_0}$ *(1) and* $\ln\frac{[PFA]}{[PFA]_0}$ *(2,3) on time.*

(□, Δ - various concentrations of the added polymer).

It was not excluded, that the deviation from regime of ideal mixing would occurred. The reason of this may be poor mixing, stipulated by adhesion of polymer to the walls of reactor, flotation of polymer on the solvent surface, arising of "dead" zone (because of imperfect design or low frequency of rotation of mixer). As it was mentioned above, the polymerization of *FA* in toluol is a heterogenic process. Therefore, presence of polymer may induce the change of intensity and degree of mixing.

For the identification the interconnection between polymer and character of mixing the series of trials was carried out, in which polymer was used as indicator. The addition of polymer in the system was performed in two modes. In first case the accumulation of polymer was proceeded in reactor, therefore the delivery of solvent was stopped after achieving by reaction of the stationary state. In this period the reactor was worked in the regime of halfperiodical process, but concentration of *PFA* was increased. After definite time the delivery of solvent was renewed and a decrease of concentration of polymer per time was observed. (Fig. 1, curve 2). In the second case after establishing in reactor of the stationary regime the definite amount of *PFA* obtained in the analogous conditions was instantly added and its concentration change per time was determined (Fig 1, curve 3). It should be noted, that independently of both, used mode and amount of used polymer, the concentration of *PFA* after definite time was again approximated to the initial stationary value.

If the added polymer is resistant, regarding to polymerization process, the decrease of its concentration can be explained by the following consideration: the total concentration of polymer, presumably, is equal to the sum of concentration of the formed and added polymers:

$$\frac{d[PFA]_{T)}}{dt} = \frac{d[PFA]_{P)}}{dt} + \frac{d[PFA]_{A)}}{dt}, \qquad (4)$$

where $[PFA]_{T)}$ is the total concentration of polymer, $[PFA]_{P)}$- added *PFA* concentration obtained via polymerization.

If polymerization proceeds independently of the added polymer under stationary conditions (which is confirmed by data of experiment), then

$$\frac{d[PFA]_{P)}}{dt} = 0 \quad \text{and} \quad \frac{d[PFA]_{T)}}{dt} = \frac{d[PFA]_{A)}}{dt} \qquad (5)$$

The variation of concentration of added polymer per time is described by the equation:

$$\frac{d[PFA]_{A)}}{dt} = [PFA]_{OA)} - \beta\,[PFA]_{A)} \qquad (6)$$

where $[PFA]_{OA}$ is the initial concentration of the added polymer. By integration of equation (6) the equation (7) is obtained which in its turn presents particular case of equation (1)

$$\ln\frac{[PFA]_{A)}}{[PFA]_{OA)}} = -\beta\, t \qquad (7)$$

As it is shown from Fig. 2 (lines 2 and 3) the alternation of concentration of the added polymer per time obeys the equation (7) which emphasizes the existence of ideal mixing. At the same time, calculated from Fig. 2 (lines 2 and 3) value of β varies within range 0.23-0.29 min^{-1} and it is in good correspondence with the real value.

Thus it should be confirmed that addition of polymer does not influence formaldehyde's polymerization and does not substentionally change mixing regime of polymerization system. The reactor response corresponds to the ideal mixing regime and even at the presence of polymer the composition of reaction mixture at any point is the same; therefore the total volume is active and as result, the mechanism process and kinetics should be the same.

I. Javakhishvili Tbilisi State University

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PHYSICAL CHEMISTRY

K.Amirkhanashvili, I.Rapoport, Academician T.Andronikashvili

Preparation of Capillary Columns for Studies of Sorbent-Sorbate Adsorption Interaction

Presented September 2, 1996

ABSTRACT. **Two types of quartz aluminized capillary columns are described. Testing data of "classical" and so-called "wide" capillary columns are given. Test chromatograms are also presented.**

Continuing the work in [1] we suggest the preparation of capillary columns.

Development of the original technology of preparation and coating of quartz capillary [2] is a novel contribution in capillary chromatography practice based on the requirements formulated by the authors in 1992. The major requirement is complete suppression of the activity of the inner surface of the capillary. For this purpose, the technology and equipment for making of quartz light pipes with an aluminium coating was taken as a principle. The peculiarity of the technology was isolation of the inner surface of the tube and stretching capillary from atmospheric air, as well as the removal of the inner layer of quartz from the tube (pipe) before extraction. High purity materials were used for coating the capillary with thin layer of aluminium and other metals which served as donors of the applied metal from the crater being melt by a laser ray (that is without a contact with the walls of other materials providing high purity of the melt and of applied layer as well as required blockage of the product from moisture). The capillary made in this way with its aluminium coating possesses striking strength and disactivated inner surface and also thermal stability over 400°C. Such a capillary with its inside diameter 210 mmc can be easily tied in a knot with a diameter of 6 mm and doesn't break for half a year.

Capillaries with inside diameter 530 mm were also prepared according to the developed technology. These capillaries are twisted in coil with a diameter of 160 mm and have unique flexibility. Test capillaries were also made with transition section from 530 mm to 210 mmc within the length of 15 m and corresponding straight section with 5 m length.

Application of traditional quartz capillary columns with polyamide coating SLPH on the capillary with aluminium coating shows an incsease in column efficiency (30%) and in the application extent (90%).

Experimental procedures

Chromatograph M-3700, Model 1 with an ionization flame detector (IFD) and adapter for application of samples was used to investigate the physico-chemical interactions in an adsorbent-sorbate system. Range of action of IFD comes from 8 to 40 (10^{-12})A, gas carried-helium.

To carry out studies on the isotherms of sorption over a wide range of sorbate concentration quartz capillary aluminized columns of two types were made: the "classical" one with the inside diameter 210 mcm and length of 30 m to investigate the effects in phase transitions of the adsorbent and discrimination of the knot of applying the sample and so-called "wide" column with the inside diameter 530 mcm, 15 m length.

Polymeric liquid-crystalline material from chloroform was applied as SLPH to the "classical" column. Average molecular mass of the polymer in the film of the inner surface of the capillary is 1.105 g/mol. Initiation of stitching was achieved by γ-radiation of the intensity of 200 rad for an hour. The thickness makes up 0.15 mcm, coefficient of capacity - 3.5, efficiency of application - 82%; column testing on heating at a rate of 5°C/min was performed according to Grob [2].

The efficiency of separation in respect to n-C_{10} amounts to 4500 thp/m.

Silicone elastomer SE-30 from acetone as SLPH was applied to the "wide" column. Application was performed in three layers to obtain the thick layer.

Peroxide of dicumol was used as an initiation of stitching. Thermal stitching is followed by washing off low molecular polymerhomologues with elastomer solvent. The thickness of the film is determined chromatographically and the efficiency of application is defined as a fraction of chromatographic capacity of the washed off film in per cent respect to the applied film. The thickness of the film is 1.4 mcm, coefficient of capacity is 12.5, the efficiency of application - 71%. Column test is performed in isothermal regime. Test chromatogram is given in Fig.1.

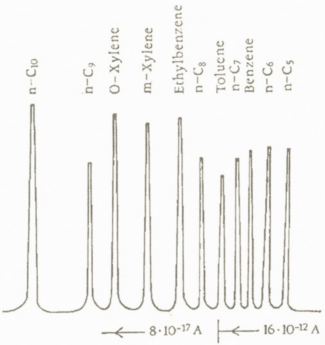

Fig. 1.

P.Melikishvili Institute of Physical
and Organic Chemistry
Georgian Academy of Sciences

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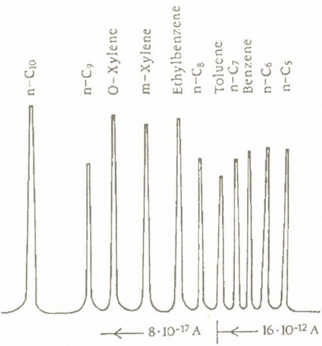

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P.Melikishvili Institute of Physical
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ELECTROCHEMISTRY

R. Kvaratskhelia, H. Kvaratskhelia

Electrochemical Behavior of I(5+) and I(7+) in the Solutions of Ammonium Salts

Presented by Corr. Member of the Academy L. Japaridze, July 15, 1996

ABSTRACT. **It has been shown that electrochemical behavior of oxyacids of I(5+) and I(7+) in the solutions of ammonium salts is characterized by the peculiarities due to appreciable proton-donor ability of NH_4^+ ions and the buffer properties of these solutions. The joint proton-donor action of NH_4^+ ions and H_2O molecules in the processes of anions electroreduction has been first described.**

The peculiarities of polarographic behavior of iodate and periodate ions in the solutions of ammonium salts have been described in [1]. The authors of this article consider the proton-donor action of NH_4^+ ions as the main reason of these peculiarities. The data obtained by us under the research of voltammetry of I(5+) and I(7+) acids-HIO_3 and H_5IO_6 [2, 3] in the various background solutions (including 0,1M $(NH_4)_2SO_4$) testify to a more complex influence of NH_4^+ ions on electrochemical behavior of I(5+) and I(7+).

The measurements were carried out by the methods of voltammetry at the rotating disk electrodes and chronovoltammetry at the stationary electrodes from high-purity Cu and Cu–Hg in the sealed cell with a pure helium atmosphere (in the case of these metals the most clear picture of the process is observed). Technique of a preparation of the electrodes for the measurements has been described in [4]. The background electrolytes used in the study - $LiClO_4$ and $(NH_4)_2SO_4$ have been twice recrystallized from a bidistillate and then have been dried for several days at 190^{0}C ($LiClO_4$) and 50^{0}C ($(NH_4)_2SO_4$). Crystalline HIO_3 has been obtained by the technique described in [5] using high-purity KIO_3, twice recrystallized from a bidistillate $BaCl_2$ and twice distilled H_2SO_4. High-purity orthoperiodic acid H_5IO_6 ("Reanal", Hungary) has been used as I(7+) containing compound. The saturated calomel electrode has been used as a reference electrode. All the measurements have been carried out at 20^{0}C.

The voltammograms of I(5+) and I(7+), shown at Fig. 1, have been registered at the Cu–Hg electrode in the HIO_3 and H_5IO_6 solutions in 0,1M $LiClO_4$ (curves 1and 2) in 0,1M $(NH_4)_2SO_4$ (curves 1' and 2') (the similar picture is observed in the case of Cu electrode). One can see from Fig. 1 that the peculiarities of electroreduction of I(5+) and I(7+) oxyanions in 0,1M $(NH_4)_2SO_4$ reveal themselves in the following experimental facts: 1) the i_d values of first waves of I(5+) and I(7+) in 0,1M $(NH_4)_2SO_4$ are appreciably less than the corresponding value for the first wave of I(5+) in 0,1M $LiClO_4$ (this wave corresponds to a $IO_3^- \rightarrow I^-$ process under the proton-donor action of H_3O^+ ions) and the total height of first two waves of I(7+) in the same background solution (here I wave corresponds to a I(7+)$\rightarrow$I(5+) process and II wave - to a IO_3^- $\rightarrow I^-$ process; both processes are also carried out under the proton-donor action of H_3O^+ ions); 2) I(7+) forms in 0,1M $(NH_4)_2SO_4$ two waves only; 3) the $E_{1/2}$ values of the last waves of I(5+) and I(7+) in 0,1M $(NH_4)_2SO_4$ are appreciably less negative than in

0.1M $LiClO_4$ (in the later case waves also correspond to a $IO_3^- \rightarrow I^-$ process, but it is carried out under the proton-donor action of water molecules); 4) the total height of the I(7+) voltammogram in 0,1M $(NH_4)_2SO_4$ (the i_d value of I(7+)) exceeds appreciably the similar value in 0,1M $LiClO_4$. First of all, it is necessary to consider first two interconnected facts. It has been shown in [2, 3] that the height of I wave of HIO_3 and the total height of first two waves of H_5IO_6 are equal to the i_d value of H_3O^+ ions. The i_d values of first waves of I(5+) and I(7+) in 0,1M $(NH_4)_2SO_4$, as it has been mentioned above, are appreciably less than the corresponding values in 0,1M $LiClO_4$. This fact shows that adding 10^{-3}M HIO_3 or H_5IO_6 to 0,1M $(NH_4)_2SO_4$ the concentration of hydrogen ions won't be equal to 10^{-3}M (as in the case of 0,1M $LiClO_4$), but will be less than this value (in the case of 10^{-3}M H_5IO_6 it is sufficient for one wave formation only and in the case of 10^{-3}M HIO_3 for the formation of I wave with a comparatively low i_d value). Comparison of the heights of the corresponding waves in 0,1M $LiClO_4$ and in 0,1M $(NH_4)_2SO_4$ helps to estimate hydrogen ions concentration in 0,1M $(NH_4)_2SO_4$ with 10^{-3}M HIO_3 (or H_5IO_6). This concentration is equal to $6{,}23 \cdot 10^{-4}$M. Decrease of hydrogen ions content is explained by the hydrolysis reaction proceeding in the solutions of ammonium salts:

$$NH_4^+ + H_2O \rightleftarrows NH_4OH + H^+ \qquad (1)$$

While adding 10^{-3}M HIO_3 (or H_5IO_6).to 0,1M $(NH_4)_2SO_4$ the part of hydrogen ions participates in the equilibrium (1). As a result their concentration decreases up to $6{,}23 \cdot 10^{-4}$ M and forms the medium with pH~3,2. This acidity maintains its value under reduction of I(5+) and I(7+) due to proceeding of reaction (1) (hence, unlike the case of 0,1M $LiClO_4$, the buffer medium with pH~3,2 and certain capacity is formed).

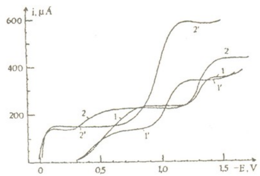

Fig. 1- The voltammograms of I(5+) a I(7+) at the amalgamated copper electrode. 980 r.p.m. 1,1'-10^{-3}M HIO_3; 2,2'-10^{-3}M H_5IO_6; 1,2-0,1M $LiClO_4$; 1',2'-0,1M $(NH_4)_2SO_4$.

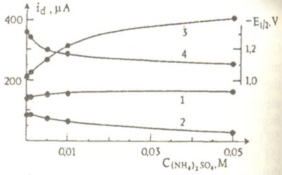

Fig. 2- The dependence of i_d and $E_{1/2}$ of I(7+) on the ammonium sulfate concentration in 0.1M $LiClO_4$. Electrode- Cu–Hg; 10^{-3}M H_5IO_6; 980 r.p.m. 1-i_d of I wave; 2- i_d of II wave; 3- i_d of III wave; 4- $E_{1/2}$ of III wave.

Due to more pronounced proton-donor ability of ammonium ions (in comparison with water) the $E_{1/2}$ values of the last waves of I(5+) and I(7+) in 0,1M $(NH_4)_2SO_4$ are appreciably less negative, than in 0,1M $LiClO_4$ (Fig. 1). It is interesting that in spite of big difference between $E_{1/2}$ values, being noticed during reduction processes of I(5+) and I(7+) under the proton-donor action of NH_4^+ ions and water molecules, under the gradual increase of the content of NH_4^+ ions in 0,1M $LiClO_4$ (Fig. 2; in the case of HIO_3 the similar change of the i_d values of I wave and the $E_{1/2}$ values of II wave is observed), separate waves corresponding to these processes are not formed: the first waves of I(5+) and I(7+) are gradually shifted to a less negative direction. This experimental fact gives us possibility to make important conclusion: the joint proton-donor action of NH_4^+ ions and H_2O molecules should be possible (that is the formation of the electrochemically active associates of anion not only with NH_4^+ ions or water molecules only is possible, but also formation of the mixed electroactive particles containing both proton donors).

It is especially important to note that the total height of the waves of I(7+) in 0.1M $(NH_4)_2SO_4$ exceeds appreciably the i_d value of I(7+) in 0.1M $LiClO_4$ (Fig. 1). The diffusion coefficient of I(7+) in 0.1M $(NH_4)_2SO_4$ (which has been calculated by us with the aid of the Levich's equation and it is equal to $6.8 \cdot 10^{-6}$ cm^2/s) is close to the D values of I(7+) in the low-acidic buffer solutions (pH~2-3) [3]. This proves the main reason of the i_d values increase of I(7+) in the solutions of ammonium salts by the fact that under the addition of H_5IO_6 in 0.1M $(NH_4)_2SO_4$ the low-acidic buffer medium is formed being the most favourable for the formation of the molecular-acidic (non-condensed) forms of I(7+) (mainly metaperiodic acid HIO_4). Thus in the solutions of ammonium salts the conditions for reaching the maximum currents of I(7+) reduction are created.

R.Agladze Institute of Inorganic Chemistry
and Electrochemistry
Georgian Academy of Sciences

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GEOLOGY

F.Maisadze

Main Lithofacies Types of Paleocene Deposits of Georgia

Presented by Academician I.Gamkrelidze, July 15, 1996

ABSTRACT. **On the territory of Georgia Paleogene deposits have been represented by three different facies: flysch, volcanogene-sedimentary and epicontinental-marine. Flysch deposits are distributed on the Southern Slope of *the Greater Caucasus (Paleocene-Eocene) and in the Adjara-Trialetian* zone (Paleocene-Lower Eocene), volcanogene-sedimentary – in the Adjara-Trialeti (Middle-Eocene and partlially Upper Eocene) and Locki-Karabakh (Middle Eocene) zones, as for the epicontinental-marine – mostly within the limits of Georgian Block (Paleocene-Oligocene).**

In Paleogene three different types of formation: flysch, epicontinental-marine and volcanogene-sedimentary took place on the territory of Georgia in different geodynamical conditions (marginal sea of the Greater Caucasus, Transcaucasian island arc, intrarc lift of Adjara-Trialeti).

Nowadays Paleogene deposits occur in all geotectonic units on the territory under consideration except zone of the Main Range of the Greater Caucasus. Their definite part is burried under the young deposits (east and west subsiding zones of the Georgian Block) or they are overlain by allochthonous formation (Mestia-Tianeti Flysch Zone).

Paleocene. On the **Southern Slope of the Greater Caucasus** rocks of this age represented by flysch and epicontinental-marine facies.

Flysch sediments are developed in Jinvali-Gombori subzone of the Mestia-Tianeti zone and are represented by Kvetera and Shakvetila suites.

Kvetera suite is built up of clastic-limestone and sandstone-aleurolitic flysch (22-50 m).

Shakvetila suite (100-260 m) is represented by sandstone-aleurolitic flysch. Pelagic (background) sediments in it are shaly argillites with scarce interlayers of limestones and as for the turbidites - arkose-quartzy sandstones and aleurolites.

Epicontinental-marine facies are developed in the eastern part of the Adler depression. Lower part of Paleocene is presented here by thin laminated limestones and variegated marls (Mikhelripshi suite), but upper one - by variegated marls (Djeopse suite). Their total thickness is 80-100 m.

Within the bounds of Kvaisa field (Varakh-kome r.), where Paleocene deposits (as well as Eocene) were stated for the first time [1] they are represented by alternation of argillaceous pelitomorphic limestones and slaty marls.

Within the limits of **Georgian Block** Paleocene deposits crop out in the west subsiding zone and are represented by **epicontinental-marine** facies, shallow marine limestones and rarely by marls. Their thickness does not exceed some tens of metres. Saretskela section is an exception - here their thickness is about 280-310 m.

As for the east subsiding zone of the Block, here Paleocene deposits are totally covered with Neogene-Quaternary molasses. According to geophysical and geological data, here Paleogene should be presented by the same facies as on the Artvin-Bolnisi block [2].

The deposits under consideration in the *Adjara-Trialetian* zone together with the Lower Eocene are represented by flysch on one hand and epicontinental-marine deposits on the other.

Flysch formations are represented by the Borjomi suite, developed on the central and east segments of the Adjara-Trialetian zone. It is built of sandstone-aleurolitic and clastic-limestone flysch. Its thickness is about 1500 m, but in separate places - 3000 m (Teleti anticline).

The fact that synchronous formations of the Borjomi suite were represented by 1000 m thick carbonate-terrigenous rocks particularly carbonate-aleurolitic argillites with interlayers of aleurolites [3] was stated by drill-holes (Chakhati anticline) in the western segment of Adjara-Trialeti (Guria **depression** is an exception). According to drillhole data these deposits apparently are of flyschoid-type, intermediate between flysch and epicontinental-marine ones.

Epicontinental-marine deposits in the zone are represented by shallow marine variegated marls and clays and their age is Paleocene - Lower Eocene as well. They crop out in a form of a narrow stripe along Surami-Gokishuri thrust and around the anticlinal structures. Their thickness does not exceed some tens of metres.

Paleocene and Lower Eocene deposits are represented by the Borjomi suite on the northern periphery (the Algeti r.basin) of *Artvin-Bolnisi Block*, but lower part of the Paleocene in Tetritskaro district is replaced by epicontinental-marine limestones and intraformational conglomerates (40-50 m) overlain by volcanogenes (350 m) of dacitic composition.

Lower Eocene. Deposits of this age are built up of flyschoidal formations in the western part of the Southern Slope of the Greater Caucasus (middle part of the Psou and Bziphi rivers), particularly by carbonate sandstones, arenaceous limestones and marls.

In the eastern part of the Southern Slope (to the east of the Liakhvi r.) both Lower and Middle Eocene form the Kvakevriskhevi suite, where in most cases it is impossible to draw a boundary between these two stratigraphic units. Lithologically Lower Eocene of the suite, built of sandstones- aleurolitic flysch is less carbonate and is characterized with dark colour, while middle Eocene part of the suite is more carbonate and consists of sandstone-aleurolitic flysch, comprising rare, 1-2 m thick interlayers of trachyandesitic tuffs.

Total thickness of Kvakevriskhevi suite is 250-350 m.

Lower Eocene is represented by variegated marls (25-55 m) of the *epicontinental-marine* Lapsta suite in the eastern part of Adler depression.

Deposits of the age under consideration on the *Georgian Block* are also represented by *epicontinental-marine* facies, i.e. by foraminiferbearing marls and limestones of low thickness.

In the *Locki-Karabakh* zone conglomerates, gravelstones, sandstones and arenceous limestones, thickness of which varies within 1.5-70 m are located under the Middle Eocene volcanogenes and are restricted to the Lower Eocene.

Middle Eocene. In the western part of the **Southern Slope of the Greater Caucasus** Middle Eocene deposits are of flyschoidal character. Definite conformity is observed in them, i.e. in the Psou r. basin they are built up noncarbonatic sandstones flyschoids, but in the Bziphi r. basin - by arkosic-greywacke sandstones and arenceous clastic limestones.

Within the limits of Ksani-Arkala parautochthone, Middle Eocene is represented by **epicontinental-marine** facies (limestones and sandstones) offen with basal formations overlying the Mesozoic rocks.

In the eastern part of Adler depression synchronous deposits are represented by Bagnari suite marls (10-45 m).

Deposits of the age under consideration on the **Georgian Block** except the territory between v.Dzegvi and Mtskheta (where volcanogene sedimentary formations are spread) are totally presented by **epicontinental-marine** facies (limestones, marls). Within the bounds of Samegrelo, **lenses** (0.15-0.7 m) of andezite and basaltic tuffs occur.

Middle Eocene formations are the most widespread in the **Adjara-Trialeti** zone. They are built of **volcanogene - sedimentary** rocks.

In different places of the folded zone various local suites and subsuites are distinguished; to state their synchronism is often difficult. The volume of the article does not allow to characterize the above-mentioned litho-stratigraphic units. They have been considered in many works and accounts. Proceedings [4], where volcanics of Adjara-Trialeti is characterized in detail are worth mentioning.

Thickness of Middle Eocene volcanogenes varies within 1000-5000 m. Their maximum thickness is observed in the western part of the northern slope of the Meskheti ridge (7000 m).

In the environs of Tbilisi the deposits under study are represented by volcanogene-terrigenic Dabakhana suite (800 m) and olistostrome ("mixlayered conglomerats") of Tbilisi (100 m).

On the **Artvin-Bolnisi Block** and in the **Locki Karabakh zone** Middle Eocene is built as well by thick (2000-2500 m) volcanogene-sedimentary rocks. In the Lower part it is presented by massive and coars-layered aglomeratic breccia of basaltic compostion, sheets of basalt and tuffs of andesite composition (Djavakheti suite), and the upper part - by layered tuffs of andesite-dacitic and liparit-dacitic composition.

Upper Eocene. On the **Southern Slope of the Greater Caucasus** facies of this age are presented by the Ildokani suite (1100-1500 m), built by sandstone-aleurolitic **flysch**. Separate bands of coarse flysch (conglomerate-breccia) are also met in the regressive upper part of the suite.

Definite conformity is observed in spatial distribution of the Upper Eocene formations from south to north direction. This is proved by the fact, that south facies are presented by flysch, but north - by flyschoids. Besides in the latter marls and limestones prevail and turbidites are met in less amount.

In the Parautochthone of Ksani-Arkala Upper Eocene deposits alongside with normal-sedimentary rocks occur olistostromes, widespread on the Southern Slope of the Greater Caucasus [5].

In the eastern part of the Adler depression deposits of this age in the lower part are built by **epicontinental-marine** Lyrolepis marls (Egrisi suite) and coloured marls (Kldiani suite); as for the upper–comparatively thick regressive Matsesta suite (350 m).

Upper Eocene is mostly presented by epicontinental-marine marls of the Egrisi and Argveta suites on the **Georgian Block**. Different facies are developed on the western part of Racha-Lechkhumi Syncline as Agvi suite built by sandy-glauconitic limestones. Interlayers of quartz-glauconitic sandstones are met in it where content of glauconite is up to 60-70%.

In the Adjara-Trialetian zone deposits under consideration are presented by the Adigeni suite, built by volcanogenes of andesite-basaltic composition and normal-sedimentary rocks (250-800 m).

Upper part of the Adigeni suite (2000 m) is located above them, presented by the same massive and coarse-detrital rocks.

Eastward, in the eastern part of the Akhaltsikhe depression, Adigeni suite is gradually replaced by sedimentary formation; as for Tbilisi environs synchronous sediments are presented by normal-sedimentary argillo-arenceous Navtlugi (50-200 m) and nummulitic (600-850 m) suites.

Oligocene. The rocks of this age (on the territory of Georgia, as in the most parts of the Crimea-Caucasus) are presented by Maikop series, the upper part of which includes Lower Miocene as well. This lowermost part is built by alternation of carbonate clays and marls (Khadumi-suite) as for the main part by noncarbonate gypceous clays and sandstones. Thickness of Oligocene part is variable and reaches its maximum in Tbilisi environs (2600 m).

On the territory under investigation different facies of Oligocene (normal-sedimentary) are noted in Adler depression (Khosta and Sochi suites), Kakheti zone (Kinti suite) and in the eastern part of Akhaltsikhe depression.

Al.Janelidze Geological Institute
Georgian Academy of Sciences

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PETROLOGY

Corr. Member of the Academy G.Zaridze, N.Tatrishvili

About the Origin of the Earth Crust

Presented July 12, 1996

ABSTRACT. **On the basis of original investigation and according to the vast literary material the following opinion is expressed in the article, that the protocrust of the Earth and protolithosphere as a whole, the age of which, exceeds 4.5-4.6 mlrd. years, had mafite, mainly basaltoid composition. Later, during protostage of the Earth self-development its sial crust was formed gradually as the results of transformation of protocrust and protolithosphere under the influence of endogenous processes.**

It's impossible to discuss the origin of the Earth crust without considering the Earth origin as a planet in the whole [1, 2].

Nowadays all cosmogonists share opinion about the fact, that stars have been originated as a result of (gas-dust clouds) our Galaxy, the Sun is among them, within its bound aries planets of the Solar system were formed. It is determined, that nebula consists of 98% of gas, mainly of hydrogen and gallium and of 2 per cent of dust particles in the form of ice: hydrates - O, C, S, N, Cl, i.e. OH, H_2O, H_3O, H_3N, HCl, C_xH_y etc.), as well as condensates of metal oxides with difficult volatilization - Mg, Al, Ca, Ti, Fe, Ni). According to many scientists estimation from 100 mlrd stars forming our Galaxy small percent should have planet systems. It is considered, that the Sun and the planets of the Solar system were formed simultaneously by condensation of initial nebula, consisting of gas and dust, being somehow "star dust". From this point of view the substance forming the Solar system should be older than the system itself [1]. Many scientists also assume simultaneity of origination of the Sun and the planets of the Solar system soon after finishing the final nucleosynthesis in our part of the Galaxy. Texture of the Earth's shell (core, mantle, crust) is considered as a result of consecutive absorption of the growing Earth of protoplanetal material of different origin and physical conditions with different contents of the warmth emanative elements [1]. In the centre of the Earth nonradioactive and nondiathermal big relict fragments of iron asteroids were mostly concentrated. Capture of particles, cosmic dust, association of atoms and secondary bodies of the growing Earth have been increased. According to this hypothesis, it is natural to assume that formation of the external shell of the Earth took place at the final stage with the substance most enriched by radioactivity. Thus, the contents of radioactive element in the protocrust of the Earth up to the depth of 1000 km reached its maximum [1]. B.G.Khlopin doesn't share the above mentioned opinion. It is considered, the nucleosynthesis in our region of the Galaxy was over 4.7 mlrd years ago, being close to the age of the Earth and coincides with the origination of the Solar system. The age of Galaxy is estimated as 15^{+5}_{-3} mlrd years.

Thus, the age of the elements is free concept, but if we estimate it with the time of finishing of the final nycleosynthesis, so it's world be 4.7 mlrd years. It should be considered, that if the nucleosynthesis within the bounds of our Galaxy continued endlessly, at the moment of coming out of the substance from the sphere of synthesis (e.g. separation of the Solar system) stable prevalence of elements would be fixed. If the nucleosynthesis was one act and lasted for a short period of time, prevalence under observation is distinguished somehow in the moment of synthesis only at the expense of time lasting till the moment of investigation. For the case, when the entire substance of the Solar system ows to sudden synthesis, about 6.6 mlrd. years are needed to obtain contemporary relation 235/398. If entire substance was formed by homogenous synthesis with exponentionally decreased intensity by the time of finishing it 4.7 mlrd years ago, then beginning of this synthesis should be considered 14.2 mlrd years ago [1]. It should be noted as well, that to draw back time of synthesis of elements more than 5 mlrd years ago, laws of nuclear physics don't allow us [1].

The age of meteorites, the Moon and the Earth is defined by the majority of the specialists within 4.5 mlrd years, on the basis, that most of the figures denoting the age of meteorites are within these bounds. The age of iron meteorites, obtained with the help of argon method, sometimes appears to be too old (up to 6-7 mlrd years). Doubt has been expressed how these data correspond to the reality [2] as these figures, as a rule exceed the age of 4.5 mlrd years adopted for the meteorites. Besides, concretion of iron minerals according to the argon and strontium methods is not more than 5 mlrd years [2].

Really the figures of K-Ar age are not simple, that is well illustrated from the presented cited authors of the table [2]. Though, together with the presence of figures of high meaning of the meteorite age there are also low figures. E.g. age distinguished according to S_m-N_d method gives the following figures: 1.34; 1.1; 1.34, but according to the R_b-S_r method - 1.55; 1.26; 3.1; 2.6; 3.7 mlrd years. Besides five figures of the age of meteorites according to R -S method show the figures considerably less than one mlrd years.

The presented approximate figures of the age of meteorites, apparently are explained by mysterious circumstances yet for the cosmochemists: insufficient knowledge of oecumenical "macro and microsituations" in which were formed small and large bodies of the Solar system. Thus, according to the prevalent number of the figures of the meteorite's age can't be judged about their acceptability.

These difficulties have clearly been formulated in the book by E.V.Sobotovich [1], who writes that history of existence of planets and parents bodies of meteorites is difficult enough. It is supposed that formation of these bodies didn't take place simultaneously. It was accompanied by heating at the expense of dissociation of shortliving and longliving radioactive elements, physico-chemical processes and possibly influence of the supernew. According to the size of these bodies heating took place with high or low speed and consequently melting and differentiation of the component substance could be elapsed in different time. Cooling of the protoplanetal bodies, their destruction and recreation could be also distinguished in time.

E.V.Sobotovich offers the following cosmochronological succession of the Solar system (isolating substance of the Solar system, formation of planets, parents' bodies of meteorites and meteorites, found and studied with studies):

1. 15^{+5}_{-4} mlrd years ago - beginning of forming of stars of the first generation, forming the first core of elements in them at the expense of S-processes (numerous nuclear processes in the stars, setting free enough quantity of neutrons, which may be seized by a kind of distributed element, giving rise to heavier cores).

2. 12^{+2}_{-2} mlrd years ago - beginning of synthesis of heavy and superheavy core (uranium, thorium etc. are among them), supernew stars of the I type of the expense of r-process (process, responsible for synthesis of core lithium, beryllium and boron, in

the highest degree not stable conditions of starry entrails that are totally burnt in thermo-nuclear reaction), beginning of the star forming of the second and the following generations;

Table

Cosmochronological succession of events causing formation of meteorites and the Earth according to E.V.Sobotovich and V.P.Semenenko

Order	Astronomical data	Data of nuclear astrophysics and cosmochemistry mlrd years	Event
1.	18.0 ± 1.8	–	Forming of "Great explosion" in the way of the Universe
2.	12.0 ± 3	15^{+5}_{-3}	Forming of Galaxi, formation of the stars of the first generation
3.	–	14.2 - 9.5	Beginning of the synthesis of heavy and superheavy nuclides (V and Th are among them in supernew of the first type on account of r-processes)
4.	–	12.0 - 5.2	Formation of the stars of the second and the third generation. Origin of the protosolar nebula and relict bodies of the Solar system
5.	–	5.0 - 4.7	Flashing of the last supernew in the environgs of protosolar nebula. Giving the last necessary for regulating impulse and "infecting of nebula" with radioactive nuclides
6.	5.0 ± 5	4.8 - 4.6	Forming of the Solar system
7.	–	4.6 ± 0.1	Forming of the Earth

3. 5.0-4.7 mlrd years ago – forming of protosolar nebula, flashing of the final supernew star in its environs "infect" of substance of nebula "sutvived" flashing of the supernew one (optical effect);

4. 4.7-4.6 mlrd years ago – consolidation embryo and bodies from mixture "infected" with radioactive and relict substance (forming of the Sun, planetes and small bodies of the Solar system of asteroid type).

5. 4.6-4.0 mlrd years ago differentiation of substance of external parts of the planets and secondary planet's bodies of meteorites, consolidated from the material enriched with radioactivity.

6. 2.5-0 mlrd years ago destruction of the parents' bodies of meteorites, existing of fragments in the face of individual small bodies [1, p.168].

The two schemes presented above (the second is in the form of the table) about the cosmochronological succession of events are principally analogous. The second scheme is rather later and somehow fuller.

In order to interpret the age of heavy bodies of the Solar system correctly, the important thing, as we consider is the utterance of E.V.Sobotovich [1] about the time their consolidation, under which he conditionally takes certain stage in the history of existence of the cosmic substance, beginning from which it exists in the face of consolidated mass. He assumes, that methods of nuclear geochronology can give a kind of meaning to the apparent age, if it's really the age and not combined influence of nuclear physical and physico-chemical processes on the obtained relation [1, p.135].

It's evident, that the Earth (the Earth' protocrust) is older than the accepted figure 4.6 mlrd years, that is stated with the rock of the core, taking out from the well of Monchegorsk on the Koli peninsula. Two interbeds (the survived remnants from total transformation of chemical "melting") are observed in the rock of peridotite and pyroxene, their argonic age showed figures 6.4-6.5 mlrd years, while the age of including the remnants rock of the intrusion is 3 mlrd years younger. E.K.Gerling explained these relations in the following: 4.5 mlrd years calculated with the help of data about the isotopes of the lead - it is not the age of the Earth but of its lithosphere of stone mantle. But the Earth as a space object might be existed earlier than springing up lithosphere in the time of foliation of its substance [3, p.28].

We share E.K.Gerling's opinion with a little precise definition. We consider, that the interbeds in the rocks of the core are protocrusts of survived fragments of the already formed Earth. It's out of the doubt, that the contemporary Earth crust is a result of processing of protocrust, mostly of mafitic (basaltoid) composition, partly of regional metasomatism, allometamorphism (metasomatism) and metasomatic anatexis.

Further, on the basis of our investigation it should be mentioned, that peridotites and pyroxenites are not always primary. That's why we assume, that probably original rocks of interbeds of Monchegorsk are rather earlier than 6.4-6.5 mlrd years.

Analogous picture has been observed in ancient formation of East-European platform and on the Baltic and the Ukraine shields, built with granitogneiss and granitoids (2.7-2.8 mlrd years), containing blocks, areas and lines of gabro-amfibolites, amfibolites and biotite-amfibolitic gneiss with the age of 3.2-3.8 mlrd years, initial rocks of which formed ancient mostly basaltoid shell of the Earth, in result of metamorphic (allomorphic or metasomatic) processing of which were formed rocks [4] containing them. Apparently the age of enclosed initial rocks is rather older than the presented figure. We assume that, processing of the protocrust proceeded not everywhere with similar intensity. Processing of Archean basaltoides is observed analogously rather later metamorphic processes.

The above mentioned opinion belong to the facts of geological category and thus it maybe taken into consideration while discussing the composition and age of the Earth crust.

High figures of the age (3.5 mlrd years) have protoplatforms built mostly of granito-gneisses and granitoids. It's known, that forming of large bodies of these rocks in later stages of the Earth is more reduced [5, p.18].

On the basis of deductive method maybe supposed that in preceeded time, more than 3.9 mlrd years ago, granitoformation took place in more considerable volumes. This process was fulfilled after finishing of extensive volcanism in the succession, that takes place in volcanogene the so called eugeosynclinals [6]. In the result of allochemical metamorphism and metasomatism under the influence of ascending fluides of definite composition, separated from the Earth mantle, crystalloshales, granito-gneiss and granitoids.

If we proceed from the hypothesis about the forming of the Earth and planets of the Solar system in the result of accumulation of the cold gas-dust clouds and take into consideration present settling state of the Venus, the reason of which might be having quite active mantle, thus it's natural to assume the presence of active mantle of the Earth as well, at the dawn of its geological development.

Mafitic composition of protocrust and protolithosphere of the Earth can be judged according to its correlation with the Moon and the Venus. Basaltoids and anorthitenorite-troctolite have been fixed on the Moon and on the Venus - division of protoplanet substance into zones; core, massif close to the surface mantle and lithosphere of basaltoid composition.

Thus, according to our imagination, protocrust of the Earth and protolithosphere as a whole, the age of which, exceeds 4.5-4.6 mlrd years, had mafite, mainly basaltoid composition. Later, during protostage of the Earth's self-development its sial crust was formed gradually in the result of transformation of protocrust and protolithosphere under the influence of endogenous processes.

A.Janelidze Geological Institute
Georgian Academy of Sciences

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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METALLURGY

Yu.Kartvelishvili, D.Mchedlishvili

Chromium Boride Aluminium Oxide and Double Boride (Titanium - Chromium) - Aluminium Oxide Composites Preparation and Study

Presented by Academician G.Gvelesiani, June 15, 1996

ABSTRACT. **The present paper deals with elaboration of new, highly efficient technological process of preparation of composites CrB_2-Al_2O_3 and $(TiCr)B_2$-$2Al_2O_3$ by alumothermal reduction of chromium chloride and boron anhydride mix or chromium chloride, titanium chloride, titanium and boron anhydride mix.**

Thermodynamics was studied. The mechanism of composites C_2B_2-Al_2O_3 and $(TiCr)B_2$-$2Al_2O_3$ was investigated. Optimal technological parameters of the process were determined.

Due to the development of new trends of modern technology the role of composites on the base of heat-proof borides has been markedly increased [1,2].

The present paper deals with elaboration of new, highly efficient technological processes of preparation of composites CrB_2-AlO_3 and $(Ti,Cr)B_2$-$2Al_2O_3$ [3] by alumothermal reduction of chromium chloride and boron anhydride mix or chromium chloride, titanium and boron anhydride mix. By means of thermodynamic calculations dependences of Gibbs function changes from temperature were determined and the possibilities of the studied reactions were established. The mechanism of preparation of composites CrB_2-Al_2O_3 and $(Ti,Cr)B_2$-$2Al_2O_3$ was investigated and optimal technological parameters of the process were determined.

While studying thermal effects of reduction processes the following materials were used as initial components: chromium chloride (99.8% $CrCl_3$) aluminium (mark AKP), titanium (carbonyl) 99.9% and boron anhydride (pure for anal.). Chromium chloride, metal-desoxidizer and titanium were taken in stoichiometric ratioes, boron anhydride in 5% excess. Components were charged into porcelain mill and mixed for three hours under the pressure of 100-150 MPa. The pressed brickets were placed in a reactor vessel and were subjected to reduction in inert medium. Reduction reaction started at 640°C, it had exothermal character and at the expense of isolated heat temperature of a reactive mixture increased up to 900-930°C.

In order to study process of preparation of composites CrB_2-$2Al_2O_3$ and $(Ti,Cr)B_2$-$2Al_2O_3$ reaction mixture was subjected to isothermal roasting in the temperature range of 800-1100°C at ageing in every 100°C for 1-5 h, and at the reduction point equals to 640°C without ageing.

The results of chemical and spectral analysis have shown, that after termination of the reduction (640°C), free boron and bound with boron chromium when obtaining CrB_2-AlO_3 equalled to 2.2%. With the increase of temperature and length of the process of homogenizing roasting, content of free boron decreased. At 1100°C and ageing for 4h content of free boron decreased to 0.20 and 0.35% respectively. Further increase of temperature and duration of the process had no practical importance.

Roentgenophase analysis of the process to obtain CrB_2-Al_2O_3 and $(Ti,Cr)B_2$-$2Al_2O_3$ after reaction of reduction (640°C) revealed boride phases of CrB_2, CrB and $(Ti,Cr)B_2$, (Ti,Cr)B as well as of aluminium oxide. With the increase of temperature up to 900°C and of ageing length up to 3-4h amount of low borides decreased, and at the approach of 1100°C and at ageing for 4h roentgenograms showed only the phases which corresponded to CrB_2-Al_2O_3 and (Ti,Cr)B_2-2Al_2O_3.

In the secondary reaction products $AlCl_3$ was obtained, which completely sublimated from the reaction mass (b.p. 180°C).

Yield of a target product amounted to minimum 95%. Composite powders in the form of CrB_2-AlO_3 are characterised by the following chemical composition: Cr - 28.8%, B_{total} -12.5%, B_{free} -0.20%, Al_2O_3 - 57.7%. Composite powders of (Ti,Cr)B_2-2Al_2O_3 are of the following chemical composition: Cr - 15.0%, B_{total} - 12.4%, B_{free} - 0.35%, Ti -13.8%, Al_2O_3 - 58.0%.

Some technological and physico-mechanical characteristics of the composite powders of CrB_2-Al_2O_3 and (Ti,Cr)B_2-2Al_2O_3 were studied. Particle size was determined on a "Coulter" device, and it was within 40-80 mxm. Maximum relative density of cold-pressed samples at the pressure of 700 MPa equalled to 64-65%, and those of hot-pressed ones: 1) for CrB_2-Al_2O_3 at 1800°C, pressure 10 MPa and ageing for 5h - equalled to 97.3%, 2) for (Ti,Cr)B_2-2Al_2O_3 at 2000°C. 15 MPa, and ageing for 5h - to 95.2%.

Investigations of cylindric samples were conducted on wear. Samples of (Φ 14 mm, L 15 mm) CrB_2, (Ti,Cr)B_2 composites CrC_2-Al_2O_3 and (Ti,Cr)B_2-2Al_2O_3, high-cutting steel P-18, electrocorundum white and tempered St-45 and IIX-15 were investigated. Tests on wear were conducted according to the following regime: rate of abrasive friction - 0.3 m/c, load on a tested sample - 2 kg. The results of calculation of the wear are given in Table 1.

Table 1

N	Sample	Weight wear of the studied samples, g		
		5min	10min	15min
1	CrB_2	0.0603	0.0161	0.0089
2	(Ti,Cr)B_2	0.0345	0.080	0.0062
3	CrB_2-Al_2O_3	0.0506	0.0125	0.0075
4	(Ti,Cr)B_2-2Al_2O_3	0.0320	0.0074	0.0053
5	P-18 (high-cutting)	0.2081	0.1252	0.1140
6	IIIX-15 (tempered)	0.2204	0.1898	0.1747
7	St-45 "-"	0.2223	0.1958	0.1832
8	Electrocorundum, white	0.2730	0.1105	0.0944

Table 1 shows that the composite material (Ti,Cr)B_2-2Al_2O_3 revealed the best indices.

F.Tavadze Institute of Metallurgy
Georgian Academy of Sciences

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METALLURGY

N.Vashakidze, G.Vashakidze

Deformation Effectiveness Determination Method for the Case of an Oval Strip Rolling In a Vertical Oval Pass

Presented by Academician I.Jordania, July, 15, 1996.

ABSTRACT. **New expressions for displaced area reduction real degree of filling of pass and effectiveness during the oval strip in a vertical oval pass of deformation are derived. These expressions can be used in designing effective roll grooving of small section and rod mills.**

The case of an oval strip rolling in a vertical oval pass is treated [1,2]. For simplification of the task we can consider oval profile as an ellipse (Fig. 1). Thus, the area of an oval strip section and a vertical oval pass are [3]:

$$\omega_1 = \frac{\pi}{4} h_1 b_1; \tag{1}$$

$$\omega_2 = \frac{\pi}{4} h_2 b_2. \tag{2}$$

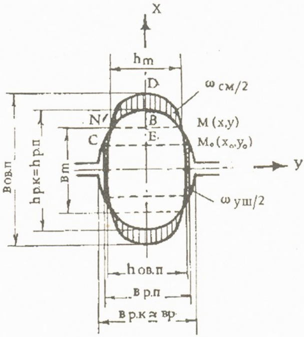

Fig. 1. The scheme of an oval strip rolling in a vertical oval pass

In these equations h_1 and b_1 represent the height and width of an oval strip (big and little axis of an initial ellipse); h_2 and b_2 represent the height and width of a final ellipse).

True index of filling for vertical oval pass is:

$$\delta = \frac{\omega_3}{\omega_2}. \tag{3}$$

Here ω_3 – represent the area of section of a vertical oval strip.

For determination of the vertical oval strip intersection area the determination of a vertical oval strip and vertical oval pass intersection point for broadening area is required (Fig.1). Vertical oval pass outline curve (ellipse) equation should be solved:

$$\frac{x_0^2}{\left(\frac{h_2}{2}\right)^2} + \frac{y_0^2}{\left(\frac{b_2}{2}\right)^2} = 1. \tag{4}$$

After solving the equation (4) relatively to x_0 coordinates, considering that $y_0 = \frac{b_3}{2}$, we receive:

$$x_0 = \frac{h_2}{2}\sqrt{1-\left(\frac{b_3}{b_2}\right)^2} \approx \frac{h_2}{2}\left[1-\frac{1}{2}\left(\frac{b_3}{b_2}\right)^2\right], \tag{5}$$

Here b_3 - represents the width of a vertical oval strip.

The area of a vertical oval strip intersection is [1]:

$$\omega_3 = 2\omega + 4x_0y_0. \tag{6}$$

The segment area is (M_0BC):

$$\omega = \frac{1}{4}h_3b_3\arccos\frac{2x_0}{h_3} - x_0y_0,$$

or

$$\omega = \frac{b_3}{4}\left\{h_3\arccos\left[\frac{h_2}{h_3}\left(1-\frac{1}{2}\left(\frac{b_3}{b_2}\right)^2\right)\right] - h_2\left[1-\frac{1}{2}\left(\frac{b_3}{b_2}\right)^2\right]\right\}. \tag{7}$$

So, finally the area of a vertical oval strip is

$$\omega_3 = \frac{b_3}{2}\left\{h_3\arccos\left[\frac{h_2}{h_3}\left(1-\frac{1}{2}\left(\frac{b_3}{b_2}\right)^2\right)\right] + h_2\left[1-\frac{1}{2}\left(\frac{b_3}{b_2}\right)^2\right]\right\}. \tag{8}$$

For the determination of a replaced area, coordinates of oval strip and vertical oval pass curves intersection point should be known (Fig.1). For this we have to solve the system of equations:

$$\left.\begin{aligned}\frac{x^2}{\left(\frac{b_1}{2}\right)^2}+\frac{y^2}{\left(\frac{h_1}{2}\right)^2}=1\\ \frac{x^2}{\left(\frac{h_2}{2}\right)^2}+\frac{y^2}{\left(\frac{b_2}{2}\right)^2}=1.\end{aligned}\right\} \tag{9}$$

After solving this system we receive:

$$b_m=2x=2\sqrt{\frac{\left(\frac{b_1}{2}\right)^2\left[1-\left(\frac{b_2}{h_1}\right)^2\right]}{1-\left(\frac{b_1}{h_2}\right)^2\left(\frac{b_2}{h_1}\right)^2}}, \tag{10}$$

$$h_m=2y=b_2\sqrt{1-\frac{\left(\frac{b_1}{h_2}\right)^2\left[1-\left(\frac{b_2}{h_2}\right)^2\right]}{1-\left(\frac{b_1}{h_2}\right)^2\left(\frac{b_2}{h_1}\right)^2}}. \tag{11}$$

Replaced area is determined by the following equation:

$$\omega'=4\left[\int_0^{h_m/2}f_1(y)dx-\int_0^{h_m/2}f_2(y)dx\right]. \tag{12}$$

But it is more convenient to use the segment area formula [3]. In this case we receive:

$$\omega'=2(\omega_{MDN}-\omega_{MBN}), \tag{13}$$

Thus, replaced area is:

$$\omega'=2\left[\frac{1}{4}h_1b_1\arccos\frac{2x}{h_1}-\frac{1}{4}h_2b_2\arccos\frac{2x}{h_2}-xy+xy\right],$$

or

$$\omega'=\frac{1}{2}\left(h_1b_1\arccos\frac{2x}{h_1}-h_2b_2\arccos\frac{2x}{h_2}\right), \tag{14}$$

$2x$ is determined by equation (10).

Finally we have:

$$\omega_1 = \frac{1}{2}\left\{ h_1 b_1 \arccos\left[\frac{2}{b_1}\sqrt{\frac{\left(\frac{b_1}{2}\right)^2\left[1-\left(\frac{b_2}{h_1}\right)^2\right]}{1-\left(\frac{b_1}{h_2}\right)^2\left(\frac{b_2}{h_1}\right)^2}}\right] - \right.$$

$$\left. -h_2 b_2 \arccos\left[\frac{2}{h_2}\sqrt{\frac{\frac{b_1}{2}\left[1-\left(\frac{b_2}{h_1}\right)^2\right]}{1-\left(\frac{b_1}{h_2}\right)^2\left(\frac{b_2}{h_1}\right)^2}}\right]\right\}. \quad (15)$$

Relative reduction and index of deformation effectiveness [4] can be determined by the formulae:

$$U = \frac{\omega'}{\omega_1}, \quad (16)$$

$$k = (1-\lambda_2)\frac{U-1}{U}, \quad (17)$$

Here λ_2-represents the stretching index in a vertical oval pass.

F.Tavadze Institute of Metallurgy
Georgian Academy of Science

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

MACHINE BUILDING SCIENCE

A.Sulamanidze

Study of Technological Possibilities of New Parameters of Electric Resistance Welding

Presented by Academician T. Loladze, June 27, 1996

ABSTRACT. **Here are presented the investigation results of the effect of new parameters such as: interaxial distance of intershifted electrodes and the second compressive force in one of near-electrode contacts on resistance welding quality, which makes the method of resistance welding proposed by the author more original.**

It is stated that shifting of electrodes brings about the useful power redistribution in separate sections of interelectrode area and regulates welding location point.

A perfect method of welding has been worked out for solution of certain problems of unipolar current microwelding, particularly to avoid plus-sign electrode working surface adhesion to the welding sample [1]. Unlike the traditional one this method has two new parameters: interaxial distance of electrodes X and second pressure force p' (Fig.1).

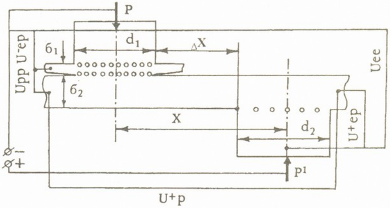

Fig.1. A flat physical model of the proposed welding method

In order to study design technological features of the proposed method and to reveal optimum conditions of welding it was necessary to exactify welding zone method for measuring electric parameter. With this purpose qualitative estimation of the process has been carried out using flat physical model (Fig.1). Model material was 0.2 mm thickness stainless steel (as well as aluminum alloy microthickness) plates. Direct current source having the following electric data: current strength (0,3-10)A, voltage (12-20)V was switched on into model. Simulation was done at different values of interaxial (or interelectrode) distances, as well as second pressure force. The obtained data showed that:

1. The increase of interelectrode distances causes decrease of current amplitude values; voltage fall amplitude values in welding zone contacts are also decreased. Interelectrode area of thick walled sample is an exception: here voltage fall is considerably increased.

2. The increase of second pressure force results in the increase of current amplitude values and rigorous fall of voltage in "positive" electrode contacts; while in other areas of welding zones voltage fall increases in accordance with current increase *(Fig.2)*.

3. *In the case* of the worked out method for calculation of total value of interelectrode zone voltage fall, the voltage fall on electrode interaxial area of the lower sample (Fig.1) should necessarily be taken into account, i.e.

$$Uee=U^{-}ep+Upp+U^{+}p+U^{+}ep$$

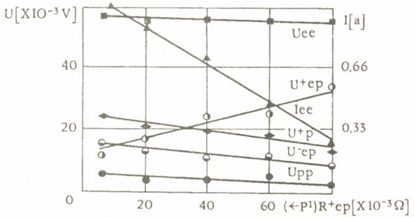

Fig.2. The effect of pressure force variation of shared "positive" electrode contact (from below) on voltage fall variation character at different sections of interelectrode area

On the next stage of the work the effect of new parameters of the mentioned method of resistance welding on seam quality was investigated. Tests were carried out on a specially assembled capacitor supply point-rolling machine, type K-51, designed at the welding department of the Moscow Bauman Technical University [2] and with K-29 type energy dosimeter. 2-6 mm diameter and 2-3 mm curving radius copper electrode of M-1 grade was used as electrode material. Grade NP-1 nickel sheet material with critical thickness ratio 6:1 was used as welding sample [3]. The carried out work proved that:

1. The increase of interaxial distance of electrodes, X, or interelectrode distance, ΔX, results in desired variation of liquid nucleus location (Fig.3.1) and, at the same time, in reduction of current amplitude value (Fig.3.2) and of liquid nucleus dimensions (Fig.3.1). It has been stated that the deviation of electrodes from alignment must occur by distance where: the first number is a constructural minimum and the second - a technological maximum. Its further increase needs welding current value increase, resulting in overheating of the lower sample on the section by overincreasing of electric power.

2. The increase of the second compressive force in positive-sign electrode contact over the desired value of welding pressure force, from 20 to 70 kg/mm, results in the decrease of voltage amplitude value in this contact from 0.2V to 0.05 V (Fig.4), which

excludes the adhesion of welding sample to electrode. At the same time welding current amplitude values are increased by 8-10%. This causes the increase of liquid nucleus dimensions 1.5 times and the decrease of electric power needed for welding.

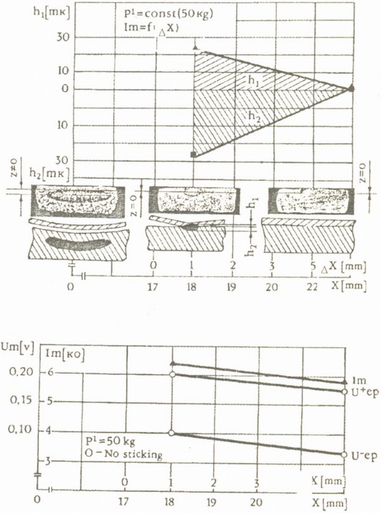

Fig.3. The effect of electrode interaxial distance on:
1. Place of liquid nucleus.
2. Amplitude values of welding currents and voltage falls in electrode contacts.

The final analysis of the present work brings us to make conclusion that

1. The simulation of the proposed perfect method of resistance welding with two new parameters allows to select electric parameters measuring bases for different sections of interelectrode zones of real welding process.
2. Its practical implementation exposes the result of electric power useful redistribution in separate sections of welding zone.
3. It allows not only to avoid adhesion of welding sample to electrode, but also to obtain normal location welded point with a high index of seamy degree for any ratio of thickness [4].

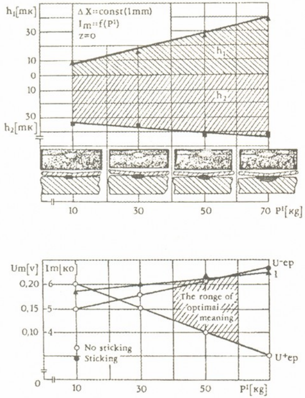

Fig.4. The effect of "Positive" (shared) electrode contact pressure on:
1. Place of liquid nucleus;
2. Amplitude values of welding current and voltage falls in electrode contacts.

Moscow Bauman Technical University Georgian Technical University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

HYDRAULIC ENGINEERING

D.Gubeladze

Determination of Free Surface Slope in Bed Flows Taking Filtration into Account

Presented by Academician O. Natishvili, November 1, 1996

ABSTRACT. **Description of hydraulic processes directly in some thin layers is a complicated problem. To draw up a logical scheme and obtain some conditional properties in microarea with the subsequent interpolation we must study the boundary conditions.**

In case of bottom permeability the following relation is considered

$$u(\eta)=(\alpha+1)\,\bar{u}\left(\frac{z+0.5d}{h}\right)^{\alpha} \tag{1}$$

When $z=0$; $u=u_0$; $z\geq 0$.

Where $u(\eta)$ is an average velocity on vertical on a given horizon; $\bar{u}$ is an average velocity of flow on vertical; d is a characteristic size of particle; α is a degree index; h is flow depth.

Nuner's relation [1] is considered. It connects degree index with the coefficient of hydraulic resistance.

$$\sqrt{\lambda}=K\alpha \tag{2}$$

Hydraulic resistance coefficient for determination is specified by some relations connected with it

$$\lambda(y)=\frac{2u_*^2(y)}{\bar{u}^2(y)}\cdot\frac{h(y)}{R(y)}, \tag{3}$$

where $\lambda(y)$ is a hydraulic resistance coefficient on vertical; y is a distance from bank to vertical; u_* is a dynamic velocity; $R(y)$ is a hydraulic radius on vertical.

Accounting integral resistance in rivers m in natural beds is calculated by the following relation

$$m=2\frac{h(y)}{R(y)}, \tag{4}$$

where m - is a bed shape.

The relation is obtained, which connects dynamic and average velocities of flow.

$$u_*=æ\,\alpha\,\bar{u} \tag{5}$$

Followed from hydraulic resistance and equation condition the following relation between æ and Karman parameter is obtained

$$K=æ\sqrt{m} \tag{6}$$

where K is empiric coefficient; æ - Karman's parameter.

Using a formula (5) the relation is obtained to calculate pulsation properties of flow:

$$\sigma_i / u_* = æ\, \alpha(a_i + b_i \sqrt{\eta}) \qquad (7)$$

where σ_i - a pulsation standard a_2 and b_i - are empiric coefficients; $(i=1,2,3)$

The spatial problem is considered for bed flow using Reinold's and continuity equations.

Going over into a system of natural coordinates i. e. if we choose a stream-line *(l)* for axis, orthogonal curved-line and cross-section stream-line *(b)* which divides flow designing into parts equal in discharge we will obtain a system of aquations approximated to N.Bernadski's equation in natural coordinates:

$$I_l = æ^2 \alpha^2 F_r - \frac{\hat{u}^2}{g}\left(\frac{1}{r_*} + \frac{1}{r_{**}}\right) \qquad (8)$$

$$I_b = F_r \qquad (9)$$

$$\frac{\partial Re}{æ} + \frac{1}{r_*} Re = 0 \qquad (10)$$

where I_l - is a surface slope along the stream-line $Fr = \hat{u}^2 / gh$; $Fr_2 = \hat{u}^2 / gr$; r - is a radius of stream-line curvature $Re = \frac{\hat{u}h}{\nu}$; $r_* = \left(\frac{\partial \theta_0}{\partial l}\right)^{-1}$ -is a radius of curved-line cross-section. $r_{**} = \left(\frac{1}{h} \cdot \frac{\partial h}{\partial l}\right)^{-1}$ - is a value characterising bottom curvature along the stream-line (curvature radius of vertical by N.Bernadski) θ_0 - is an angle between stream-line and axis $x(\theta_0 \approx \partial l / \partial x)$ [2].

As follows from a formula (8) in case of smooth flow, when r_* and r_{**} are great enough the following relation is correct

$$I_l = æ^2 \alpha^2 F_r \qquad (11)$$

The relations (8) and (9) can be used to design a flow by the method of N. Bernadski, then the basic design formula will be as follows:

$$K_{i+1} = K_i \left(\frac{h_{i+1}}{h_i}\right)^{3/2} \frac{\alpha_i}{\alpha_{i+1}} \qquad (12)$$

where $K_i = \frac{\Delta l}{1/2(\Delta b_{i+1} + \Delta b_j)}$ -is a coefficient characteristic for relation of cross Δl and longitudinal Δb square size. i, j - is a stream-line number.

Using the relations (8), (9), (10) we will obtain the following equation

$$\frac{æ}{g \Omega I_l} \int_0^h \alpha^2 \hat{u}^2 \, dy = 1 \qquad (13)$$

where Ω is a free cross-sectional area.

Relation (13) is used to determine free surface average slope on edge and calculate in rivers and channels as well.

As experimental data show all the investigations are reduced to determine α and $\hat{u}$ properties of the main parameter of bed flows. Filtration problem in experiment can be solved by measuring a discharge and average velocity.

Georgian State Agrarian University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

V.Debolsky, D.Gubeladze

Statistical Description of Turbulent Motion in Bed Flows

Presented by Academician O.Natishvili, November 1, 1996

ABSTRACT. **This work reports the description of turbulent motion in bed flows and turbulent processes in rivers. Such properties as distribution of averaged velocities, tensor of Reinolds stress, law of distribution of pulsation velocities, willer and Lagrange integral scales relation are considered.**

To determine a nature of flow velocities distribution a pulsation standars of proportional correlation coefficient $r = \overline{u'w'}$ is proposed and a pulsation itself:

$$\overline{u'w'} = -u_*^2 (1-\eta) \qquad (1)$$

where $u'w'$ are components of cross and longitudinal vertical velocities.

u_* is a dynamic velocity; σ_u and σ_w - a standard of cross and longitudinal vertical velocities.

All the three components of pulsation velocity are described as follows:

$$\frac{\sigma_i}{u_*} = a_i + b_i\sqrt{\eta} \qquad (2)$$

where σ_i is i velocity's standard; a_i, b_i - are empiric coefficients (i = 1, 2, 3).

Taking into account [1] and [2] formulae a distribution of correlation coefficients along the full length of flow is obtained:

$$r_{u,w} = -(1-\eta)/(a_1 + b_1\sqrt{\eta})(a_2 + b_2\sqrt{\eta}) \qquad (3)$$

Autocorrelation function of measurements decreases monotonously with the increase of argument and can be approximated with sufficient accuracy by the following relations obtained experimentally

$$r_i(T) = \exp(-T/(Q_i))_w \qquad (4)$$

where $T = \frac{\overline{u}\,t}{h}$; $(\theta_i)_N = \int_0^\infty r_i(T)\,dT$ is a correlation radius or Willer's integral turbulence temporary scale.

Cross, longitudinal and vertical scales, components of pulsation velocity for isotropic turbulence are equal 2:1:1 [1].

Solving the problem of spreading passive admixture in a plane flow is reduced practically to calculation of exchange coefficient. Thus let's use Tailor's representation:

$$\varepsilon_i = \sigma_i^2 (\theta_i)_L \qquad (5)$$

As a first approximation - $\sigma_i \approx \alpha_i(\eta)u_*$; $(\theta_i)_L \approx \beta_i(\eta)\frac{h}{\overline{u}}$; α_i, β_i - some function η; $\eta = z/h$.

Admixture distribution in flow planning can be calculated with the following relation

$$P_0(x, y) = (2\pi\sigma_x\sigma_y)^{-1}\exp-\left[\frac{(x-\bar{x}^2)}{2\sigma x^2}+\frac{y^2}{2\sigma y^2}\right] \tag{6}$$

where $\bar{x} = \bar{u}\, t;\ y = 0$.

Dispersion σ_x^2 and σ_y^2 will be found by a formula $\sigma_x^2 = 2\varepsilon_1 t;\ \sigma_y^2 = 2\varepsilon_2 t;\ \sigma_x,\ \sigma_y$ is a standard of particle deviation from x and y axes; t is a time

ε_1 and ε_2 are determined by the relation:

$$\varepsilon_1 = \psi_i\,\lambda/2\ \bar{u}\,h = C_i\,\bar{u}\,h, \tag{7}$$

$$\varepsilon_2 = \psi_i\sqrt{\lambda}\,/2\ u_* h = K_i\,u_* h, \tag{8}$$

where ψ_i, K_i, C_i coefficients are obtained on the basis of experimental data.

The modelling of turbulent diffusion is carried out using the method of wandering particles. The scheme chosen for modelling is maximally approximated to the experiments [2,3].

Particles trajectory is calculated by successive movement from point $/x_i,\ y_i/$ to $/x_{i+1}$, $y_{i+1}/$ point in a small interval of Δt time.

$$x_{i+1} = x_i + \bar{u}_i\,\Delta t + \sigma_u\xi_{i+1}\Delta t\cos 2\pi\varphi_{i+1}, \tag{9}$$

$$y_{i+1} = \tilde{y} + \sigma_v\xi_{i+1}\sin 2\pi\xi_{i+1},$$

where ξ is a normal random value by correlation given in advance; φ is a value distributed equally; $\tilde{y}$ is determined from condition $u(x,\ y)$ = const; $\xi_{i+1} = \xi_i\, r(\Delta t) + \alpha_{i+1}\sqrt{1+r^2(\Delta t)}$; $r(\Delta t) = e^{-\Delta t/\theta_1}$; ξ_{i+1} is a random value of normal numbers; $r(\Delta t)$ is a correlation given in advance, Δt step of calculation in time.

On the basis of experimental data the following relation is obtained:

$$\frac{\sigma_c^2}{\sigma^2} = \frac{1}{a}\left[1+\frac{1}{2a}\left(1-e^{-2a}\right)\right] \tag{10}$$

Results of statistical modelling with the proposed algorithm coincide fairly with the results of measurement.

On the basis of experimental data the relation is obtained determining a mean distance of particles having a fall diameter from bottom in time T.

$$\bar{\eta}_{\bar{\omega}_0} = \bar{\eta}_0 - \frac{\omega_0 T}{\bar{u}}, \tag{11}$$

where $\bar{\eta}_0$ is a mean distance of neutral buoyant particles from bottom ($\omega_0 = 0$)

ω_0 is particle's fall diameter.

The relation (11) is verified by the results of statistic modelling. Using Navie-Stocks equation the following relation is obtained to assess a critical velocity on the bottom

$$u_{c.v.} = \frac{\omega_0}{2C_0}, \tag{12}$$

where C_0 is a coefficient of resistance towards displacement of particles. To determine a fall diameter the following relation is used

$$\omega_0 = 1,15\sqrt{\left(\frac{\rho_s - \rho_w}{\rho_w}\right)\frac{g \cdot d}{C \cdot K_0}} \tag{13}$$

where C is a coefficient of sphere streamline resistance determined for different Reinolds number calculated from particle's characteristical size $Re_d = \frac{\overline{\omega}_0 d}{\nu}$; g is acceleration of gravity; ρ_s , ρ_w - respectively density of water and sediment; K_0 is a coefficient of particle's shape.

Critical bottom velocity of flow P_0 must be determined taking into account a necessity of probability of particles displacement

$$u_{c.b.} = u_{c.v.}\left[\ln\left(l + \frac{p_0}{0,3} - l\right)\right] \tag{14}$$

where l is a natural logarithmic basis.

If P_0 value is given we can be guided by considerations: if $P_0 = 0$ particles is not displaced, but when $P_0 = 1$ the particles begin to separate from bottom. The obtained relation coincide well enough with the data calculated directly.

Thus we can conclude the obtained statistical model can be widely used to describe turbulent motion in bed flows.

Georgian State Agrarian University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

AUTOMATIC CONTROL AND COMPUTER ENGINEERING

I.Mikadze, R.Kakubava

Time-Characteristics Analysis of the Queueing System with Cumulative Faults and General Erlang Arrivals

Presented by Academician V.Chichinadze, October 10, 1996

ABSTRACT. **A queueing system where service medium is a subsystem with many operating states has been studied. Probabilistic characteristics of the response time and the sojourn-time for the transient state are obtained.**

A general Erlang flow of calls with parameters $\lambda_0, \lambda_1, ..., \lambda_{l-1}$ enters a queueing system. The server is subjected to permanent faults which are detected at the moment of their inception, but it keeps to function till the quantity of the faults is less than m, and after m faults is repairing. The repair time is a random variable with distribution function $G(u) = \int_0^u g(v)\,dv$. The faults rate in serviceability state of server is equal to α_i.

The waiting time and the queue length in the system are unbounded.

It can be inferred from the foregoing consideration that the server can be in one of m serviceability states at the beginning of processing for each call; the same states are possible at the end of each service.

To describe the behavior of the queueing system during the servicing of the calls we introduce the notations:

$$H_{ij}(u) = \int_0^u h_{ij}(v)\,dv;\ H_i(u) = \sum_{j=0}^{m-1} H_{ij}(u)\ ;\ i,j = \overline{0, m-1}.$$

Here $h_{ij}(u)du = P$ {the servicing of a call begun by the server in i state will be completed in j state during the time interval $(u, u+du)$}.

No other assumptions are made with respect to the system in the busy state, because all possible variants (modes) of operation of the unreliable server affect only the form of the functions $H_{ij}(u)$. We regard these functions initially as the primary characteristics of the system, setting as our objective to investigate its time-characteristics. The operation of the considered system is described by the following probabilities:

1) $R_i^{(k)}(t) = P$ {there are no calls in the system at the moment t; the server is in the i serviceability state; an entering call is in the k phase of the Erlang incoming scheme};

2) $q_i^{(k)}(t, u)du = P$ {the server processes the call or it is repairing at the moment t; the entering call is in k phase of the Erlang incoming scheme; the server will be able to process the call after time, the length of which lies in the interval $(u, u+du)$ and it will be in i serviceability state at this moment}; $i = \overline{0, m-1}$; $k = \overline{0, l-1}$.

Using the standard probabilistic arguments on the possible changes of the system state in the time interval $(t,\ t+\Delta t)$ and passing to the limit as $\Delta t \to 0$, we obtain the following difference-differential equations for the probabilities defined above:

$$\partial q_i^{(0)}(t,u) \,|\, \partial t - \partial q_i^{(0)}(t,u) \,|\, \partial u = -\lambda_0 q_i^{(0)}(t,u) + \delta_{i0}\alpha_{m-1} R_{m-1}^{(0)}(t) g(u) +$$

$$+ \sum_{v=0}^{m-1} \lambda_{l-1} \int_0^u q_v^{(l-1)}(t,v) h_{vi}(u-v)dv + \lambda_{l-1} \sum_{v=0}^{m-1} R_v^{l-1}(t) h_{vi}(u); \qquad (1)$$

$$\partial q_i^{(k)}(t,u) \,|\, \partial t - \partial q_i^{(k)}(t,u) \,|\, \partial u = -\lambda_k q_i^{(k)}(t,u) + \delta_{i0}\alpha_{m-1} R_i^{(k)}(t) g(u) +$$

$$+ (1-\delta_{k0})\lambda_{k-1} q_i^{(k-1)}(t,u),\ k = \overline{1, l-1}; \qquad (2)$$

$$d R_i^{(k)}(t) \,|\, dt = -(\alpha_i + \lambda_k) R_i^{(k)}(t) + (1-\delta_{i0})\alpha_{i-1} R_{i-1}^{(k)}(t) +$$

$$(1-\delta_{k0})\lambda_{k-1} R_i^{(k)}(t) + q_i^{(k)}(t, 0),\ k = \overline{0, l-1}. \qquad (3)$$

Here δ_{ij} is the Kronecker delta:

$$\delta_{ij} = \begin{cases} 1, & \text{for } i = j \\ 0, & \text{for } i \neq j \end{cases}.$$

We assume that initially the server is in the "o" serviceability state, there are no calls in the system and an entering call is in the "o" phase of the Erlang incoming scheme, i.e.

$$R_i^k(0) = \delta_{io}\delta_{ko};\ q_i^k(0, u) = 0;\ i = \overline{0, m-1};\ k = \overline{0, l-1}.$$

The Lapalace-Stieltjes transform (1-3) by both arguments (first by t and then by u) we obtain after simple manipulations:

$$(s - \omega + \lambda_0)\overline{\overline{q}}_i^{(0)}(s,\omega) - \sum_{v=0}^{m-1} \lambda_{l-1}\, \overline{h}_{vi}(\omega)\, \overline{\overline{q}}_v^{(l-1)}(s,\omega) = -\overline{q}_i^{(0)}(s, 0) +$$

$$+ \delta_{i0}\alpha_{m-1} \overline{R}_{m-1}^{(0)}(s)\, \overline{g}(\omega) + \lambda_{l-1} \sum_{v=0}^{m-1} \overline{h}_{vi}(\omega)\, \overline{R}_v^{(l-1)}(s); \qquad (4)$$

$$(s - \omega + \lambda_k)\overline{\overline{q}}_i^{(k)}(s,\omega) = -\overline{q}_i^{(k)}(s, 0) + \delta_{i0}\alpha_{m-1} \overline{R}_{m-1}^{(k)}(s)\, \overline{g}(\omega) +$$

$$+ (1-\delta_{k0})\lambda_{k-1} \overline{\overline{q}}_i^{(k-1)}(s,\omega),\ k = \overline{1, l-1}; \qquad (5)$$

$$(s + \alpha_i + \lambda_k)\, \overline{R}_i^{(k)}(s) = \delta_{io}\delta_{ko} + (1-\delta_{io})\alpha_{i-1} \overline{R}_{i-1}^{(k)}(s) + (1-\delta_{ko})\lambda_{k-1} \overline{R}_i^{(k)}(s) +$$

$$+ \overline{q}_i^{(k)}(s, 0),\ k = \overline{0, l-1},\ i = \overline{0, m-1}. \qquad (6)$$

Solving the equation (6) with respect to $\overline{q}_i^{(k)}(s, 0)$, putting it in (5) and taking into account (4) we obtain after some transformations

$$a_k \overline{\overline{q}}_i^{(k)}(s,\omega) = \lambda_{k-1} \overline{\overline{q}}_i^{(k-1)}(s,\omega) + b_i^{(k)}(s,\omega),\ k = \overline{1, l-1},\ i = \overline{0, m-1}; \qquad (7)$$

$$a_k \overline{\overline{q}}_i^{(k)}(s,\omega) = \left[\sum_{c=1}^{k} b_i^{(c)}(s,\omega) \prod_{n=c}^{k-1} \lambda_n / a_n \right] + \left[\lambda_0 \prod_{n=1}^{k-1} \lambda_n / a_n \right] \overline{\overline{q}}_i^{(0)}(s,\omega); \qquad (8)$$

Here

$$a_k = s - \omega + \lambda_k;\ b_i^{(k)}(s,\omega) = \delta_{i0}\delta_{k0} - (s + \alpha_i + \lambda_k)\, \overline{R}_i^{(k)}(s) +$$

$$+ \delta_{i0}\alpha_{m-1}\, \overline{g}(\omega)\, \overline{R}_{m-1}^{(k)}(s) + (1-\delta_{i0})\alpha_{i-1} \overline{R}_{i-1}^{(k)}(s) + (1-\delta_{ko})\lambda_{k-1} \overline{R}_i^{(k-1)}(s)$$

$$[\tilde{a} - \tilde{\lambda}\,\overline{h}_{ii}(\omega)]\overline{\overline{q}}_i^{(0)}(s,\omega) - \tilde{\lambda}\sum_{\substack{v=0\\ v\neq i}}^{m-1}\overline{h}_{vi}(\omega)\overline{\overline{q}}_v^{(0)}(s,\omega) = B_i(s,\omega),\ i=\overline{0,m-1} \qquad (9)$$

The expressions $\tilde{a}$, $\tilde{\lambda}$ for and $B_i(s,\omega)$ are obtained from (4-6); at the same time the $B_i(s,\omega)$ is the linear function from $R_i^k(s)$, $i=\overline{0,m-1}$, $k=\overline{0,l-1}$.

Solving the system of equations (9) with respect to $\overline{\overline{q}}_i^{(0)}(s,\omega)$, $i=\overline{0,m-1}$, we obtain

$$\overline{\overline{q}}_i^{(0)}(s,\omega) = \Delta_m^{(i)}(s,\omega) \big/ \Delta_m(s,\omega).$$

The Laplace-Stieltjes transforms $\overline{\overline{q}}_i^{(0)}(s,\omega)$ are analytical functions of ω in Re(ω) >0 for Re(s) >0 and therefore the zeros of the denominator should coincide with the zeros of the numerator. In one's turn the determinant has m single zeros with respect to [2,3].

If the solutions of equation $\Delta_m(s,\omega)=0$ with respect to ω are denoted by $\omega_k=\omega_k(s)$ ($k=\overline{1,ml}$) then it is clearly necessary that

$$\Delta_m^i(\omega_k s) = 0,\ i=\overline{0,m-1};\ k=\overline{1,ml}. \qquad (10)$$

The system of equations (10) will be used to determine the unknowns $\overline{R}_i^{(k)}(s)$ ($i=\overline{0,m-1}$, $k=\overline{0,l-1}$). This requires ml equations. The total number of equations in (10) is $m^2 l$, but it easy to verify that only the ml equations obtained for one of m values of i are independent. In particular, in order to determine the unknowns $\overline{R}_i^{(k)}(s)$, we use the system

$$\Delta_m^{(0)}(s,\omega_k) = 0,\ k=\overline{1,ml}. \qquad (11)$$

Developing the determinant $\Delta_m^{(0)}(s,\omega_k)$ by the first column, we obtain the system on ml equations with respect to unknowns $R_i^{(k)}(s)$, we obtain

$$\sum_{v=0}^{m-1}\sum_{c=0}^{l-1} b_{vc}(s,\omega_k)\,\overline{R}_v^{(c)}(s) = -\Delta_{m0}^{(0)}(s,\omega_k),\quad k=\overline{1,ml}. \qquad (12)$$

It is easy to determine the parameters of this system from the (11).

We are acting hereafter following this algorithm:

1. we solve the system (12) and we find the expressions for $\overline{R}_i^{(k)}(s)$, $B_i(s,\omega)$ and $b_i^{(k)}(s,\omega)$;
2. we substitute the expressions for $B_i(s,\omega)$ in (9);
3. we solve the obtained system of equations and we find $\overline{\overline{q}}_i^{(0)}(s,\omega)$
4. we find the expressions for $\overline{\overline{q}}_i^{(0)}(s,\omega)$ from (8); ($i=\overline{0,m-1}$, $k=\overline{0,l-1}$.

Caring out this algorithm we will be able to make the following analysis.

We denote by $W(t)$ the virtual waiting-time at the moment t. $Q(t,u)=P[W(t)<u]$ is the waiting-time distribution.

It is easy to obtain

$$Q(t, u) = \sum_{i=0}^{m-1}\sum_{k=0}^{l-1}\left[R_i^{(k)}(t) + \int_0^u q_i^{(k)}(t, v)dv\right];$$

$$\overline{\overline{Q}}(s, \omega) = \frac{1}{\omega}\sum_{i=0}^{m-1}\sum_{k=0}^{l-1}\left[\overline{R}_i^{(k)}(s) + \overline{\overline{q}}_i^{(k)}(s, \omega)\right].$$

The expression for $\overline{\overline{V}}(s,\omega)(V(t,u))$ is the distribution for sojourn time is obtained analogously:

$$\overline{\overline{V}}(s,\omega) = \sum_{i=0}^{m-1}\sum_{k=0}^{l-1}\left[\overline{R}_i^{(k)}(s) + \overline{\overline{q}}_i^{(k)}(s,\omega)\right]\overline{h}_i(\omega).$$

Using the expressions $\overline{\overline{Q}}(s, \omega)$ and $\overline{\overline{V}}(s,\omega)$ it can obtain all time-characteristics of the investigated queueing system.

Georgian Technical University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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O.Shainidze

About New Kinds of Adjaria Microflora

Presented by Corr.-Member of the Academy G.Nakhutsrishvili, October 21, 1996.

ABSTRACT. **Systematic characteristics of 6 kinds of fungi new for science found in microflora of Adjaria are presented.**

Studying microflora of Adjaria since 1976 - 1995 6 kinds of fungi were revealed to be new for science. Here we are citing systematic characteristics of these kinds.

1. *Pestalotia machilusiae* Shainidze.

Spots are large, ellipsoidal, greyish-brown, surrounded by dark brown border. Pustules are scattered on the upper side of the spots, the wall is thin, dark-brown or blackish 155-272μ in diameter *Conidium* bearer is placed in one line, colourless, cylindrical, identical in sizes 7.5×1.7μ. *Conidium* is almost elliptical, spindle-shaped, 5 cells, 23.5-34.2×5-6μ, 3 middle green-olive cells. on the top end 4 thread-like colourless cilia 13-17×0.8μ, short colourless stem 7-8×0.9μ.

On leaves *Machilus thunbergii* Sieb. et Zuk., Batumi Botanic Garden, 11.07.1980, 19.09.1994.

2. *Pestalotia ternstroemiae* Shainidze.

Spots are fine of irregular shape, dark brown surrounded by black border. Pustules are numerous, situated in groups on the upper side of the spots, yellow, 390×148μ in diameter. Conidium bearers are short, hardly seen. Conidiums are spindle-shaped, 6 celled, 23.5-34.2×5-6μ with 4 middle brown-black cells (3-rd cell with transversed partition), with colourless upper and lower cells, on the top end with 3 colourless cilla, 19-21×1.1-1.4μ, short stem 7.5-9×1.2μ.

On leaves of *Ternstroemia gimnanthora* Wight et Arn., Batumi Botanic Garden, 19.08.1989, 26.09.1993.

3. *Phomopsis nandinae* Shainidze.

Spots on thin branches are large, surrounded with black border. Pycnidiums are numerous in groups, scattered on spots, spherical, elliptical, black, 215-325μ in diameter, having 2 kinds of spores.A-spores are colourless, oblong, spindle shaped, ellipse shaped, with one or two drops of oil, monocelled, 8.4-12.5×2.6-3.2μ B-spores thread -like, long a little, bent at the end, 15.3-19×1.2 -1.8μ.

On the branches *Nandina domestica Thunb.*, Batumi Botanic Garden 20.05.1995.

4. *Macrophoma wisteriae* Shainidze.

Pycnidiums are numerous, slightly round, spherical scattered, covered with epidermis sometimes interrupted protruding over epidermis with porusa with slightly brown-black wall, 230-310μ in diameter. Conidium bearers are colourless, oblong, cylindrical, with almost pointed ends, 17.2-23.5×2.2-2.8μ. Conidiums are colourless, grainy, spindle-shaped, sometimes ellipsoidal on sides, 18-26.4×6.5-8.8μ.

On branches *Wisteria sinensis* (Sims.) Swet, Batumi Botanic Garden, 11.09.1988; Batumi Maritime Park, 27.08.1989.

5. *Ascochyta cinnamomumiae* Shainidze.

Numerous pycnidiums on the lower side of the leaves and on thin branches scattered either in groups half immersed under epidermis or protruding over it, slightly round, ball-type or slightly round, pear-shaped with dark brown-black walls, slightly round porusa, thin dark cell 220-270μ. Conidiums are numerous cylindrical, slightly ellipsoidal, straight, or slightly bent, colourless, double-celled, among cells one side is clearly expressed, 11.5-19.5×4.7-5.2μ.

On leaves *Cinnamomum glanduliferum* (Wall.) Weiss., Batumi, 16.08.1992.

6. *Coniothyrium neolitseae* Shainidze.

Pycnidiums are in abundance, in groups and singular in the form of black dots, covered with epidermis with thin parenchymal cell, dark brown wall, round, 110-175μ in diameter. Conidium bearers are clearly expressed, cylindrical, slightly pointed at the ends, bent, seldom straight, colourless, 17.5-19.2×1.2μ. Conidiums are numerous, slightly round ellipsoidal, monocelled, slightly yellowish, thin, 2-2.5μ.

On the branches *Neolitsea sericea* (Blume) Koid., Batumi Botanic Garden, 22.05.1990, 13.09.1955.

Geeorgian Academy of Sciences
N.Ketskhoveli Institute of Botany

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №2, 1996

HUMAN AND ANIMAL PHYSIOLOGY

M.Chikvaidze, Z.Nanobashvili, G.Bekaia

Effect of Repeated Afferent Stimulations on Responsive Reactions of Neocortical Polymodal Neurons

Presented by Corr. Member of the Academy V.Mosidze, August 6, 1996

ABSTRACT. **The influence of various peripheral repetitive stimulation of neurons in the dorsal suprasylvain gyrus was studied in adult cats immobilized by myorelaxants. Habituation of responses of polymodal neurons in the dorsal suprasylvian gyrus was found to be selective. Selective habituation of responses in the dorsal suprasylvian gyres indicates the occurrence of diverse functional changes in the convergent afferent pathways.**

Coordinated intraaction between neurons interrelated by invariant connections appears to be a basis for normal activity of any region of CNS. During different behaviors stable relations between the neurons may be changed under the influence of some factors, i.e. invariant connections and correspondingly synapses undergo structural-functional changes [1-7].

In order to characterize the elements in the neural circles and to reveal their functional possibilities it is especially reasonable to use prolonged repeated stimuli evoking dynamic changes in the activity of neural cells [1-7].

It should be also noted that changes in the activity of neurons of various parts of the nervous system under the influence of prolonged afferent stimulations are different. These changes differ even in the different cells of the same structure [6, 7]. As the coordinated action of polymodal neurons is especially important in the integrative action of the nervous system, it was interesting to study the structures, where neurons responded to various afferent stimulations.

The study of neurons of association field in general and of parietal association field in particular, appears to be of great interest, because: a)neurons of above-said field are polymodal [8,9]; b)reactions of neurons in this field to afferent stimulations of different duration have not been studied.

The study was carried out on unanesthetized, curarized cats under the conditions of acute experiments. Light flashes, clicks and electrical stimulation of contra- and ipsilateral limbs were applied as peripheral stimulation. The pupil dilatation was observed by 0.1% atropin solution. Neural activity of posterior suprasylvian gyrus was registered by means of glass capillaries filled with 3 M sodium citrate.

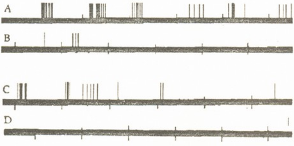

Fig.1. Changes in neural activity of parietal association field at light flashes (A,B) and sound repeated stimulations (C,D). Calibration: 250 ms, 200 mkV.

Changes in reactions of polysensory neuron of posterior suprasylvian gyrus at light flashes (A,B) and sound stimuli (C,D) are shown on Fig.1. As it is shown, after several repetitions of light flashes, the blocking of evoked discharges was observed. The neuron registered on the first and the second light flashes generated burst of discharges (10-12 spike potential). Beginning from the third light flash a progressive decrease

(Fig.1, A) in evoked discharges and their total disappearance (Fig.1, B) were observed. After the eighth stimulation responses of the neuron did not appear any more. During the restoration of responsive reactions by stimuli. with changed characteristics the phenomenon of habituation was observed again. It should be emphasized that restored neural discharges at light flashes undergo the habituation more quickly, i.e. the "potentiation" of habituation took place [1].

The time necessary for decrease of reactions in the registered neuron at sound stimuli was shorter. It has been shown on Fig.1, that after using the first two sound stimuli (C), the decrease of number of discharges up to their total disappearance (D) was observed. In order to ascertain the phenomenon of habituation we used the same criteria, as in case of light flashes.

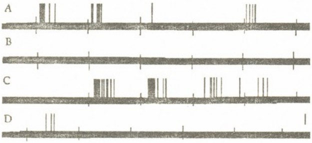

Fig.2. The same, as in Fig.1. Detailed explanation in the text.

The effect of habituation evoked in neuron by the stimulation of one afferent system on the responsive reactions induced by the activation of the other afferent system is shown in the following figures. Fig.2 shows the disappearance of reactions after repeated sound stimulation (A) just as in Fig.1. On this background light stimulation (C,D) conditioned dehabituation in the registered neuron, on the one hand, and a total disappearance (Fig.2,D) of reactions evoked at visual stimuli (after 4-5 light flashes), on the other hand.

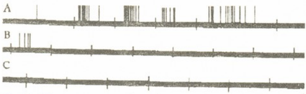

Fig.3. The same, as in Figures 1 and 2. Detailed explanation in the text.

The effect of habituation in the same neuron, evoked by light flash on the reactions evoked by sound stimulation was shown in the following figure: reactions registered at light stimulation were inhibited after 4-5 light flashes (Fig.3, A,B). It is noteworthy that the neuron registered on this background, does not generate discharges (C) to the sound stimulation. So, habituation evoked by light, conditioned habituation evoked by sound stimuli.

From the data obtained it is suggested that habituation in polymodal neurons of parietal association cortex reveals the selectivity. The mentioned changes in spike activity should indicate the different shifts occurring in afferent pathways, convergating on the registered neuron of the cortex. Besides, these changes should be conditioned by distinct mechanisms of habituation evoked by different afferent pathways [1,2, 4,5].

I.Beritashvili Institute of Physiology
Georgian Academy of Science

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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BIOPHYSICS

E.Chikvaidze, I.Kirikashvili

The Study of Complex of Albumin with Bilirubin in the Presence of Mn^{2+} Ions

Presented by Academician M.Zaalishvili, May 14, 1996

ABSTRACT. **The concentration of free bilirubin in plasma and its increase causes hepatitis.**

The dependence of the stability of triple complex of human serum albumin with bilirubin and Mn^{2+} ions on pH of solution using electronic paramagnit resonance (EPR) and the electronic spectroscopic method is studied.

Bilirubin neurotoxic pigment, which is abundant in organism as a product of a degradation of hemoglobin, makes the non–covalently bond with serum albumin in order to form its gluconorisation in the liver and to isolate it from the organism. Thus, albumin functions as a biological buffer against encelophatia. The disturbance of above mentioned complex stability causes an increase of concentration of free toxic bilirubin in blood of a heavy disease – hepatitis. Normally the concentration of free bilirubin in plasma is $(5-10)\cdot 10^{-6}$ M, its increase up to $2\cdot 10^{-5}$ M causes hepatitis, but its increase up to $(3-4)\cdot 10^{-4}$ M – is lethal.

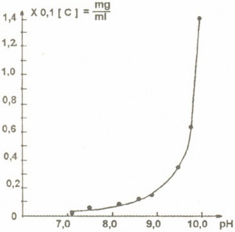

Fig.1. On the pH dependence of solubility of bilirubin in water $[C]-\frac{mg}{ml}$ the concentration of dissolved bilirubin.

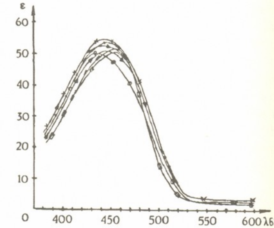

Fig.2. On the pH dependence of absorption spectrum of free bilirubin and of bilirubin–albumin complex.
▶ ▶ – pH=7.0; × × – pH=7.2;
▷ ▷ – pH=7.4; ● ● – pH=7.6
– pH=7.8; – pH=8.0
○ ○ – free bilirubin.

We studied the dependence of the stability of triple complex of human serum albumin with bilirubin and Mn^{2+} ions on pH of solution using electronic paramagnit resonance (EPR) and the electronic spectroscopic method.

The solubility of free bilirubin in water greatly depends upon pH of the solution (Fig.1). In the case of neutral values of pH, bilirubin almost doesn't dissolve in water. The solubility of bilirubin in plasma depends on its joining with albumin. In normal plasma the bilirubin solubility is 100 times and in some pathological conditions up to 5000 times more than its solubility in water [1].

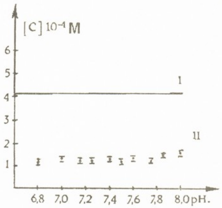

Fig.3. On the pH dependence of absorption spectrum pf albumin–Mn^{2+}–bilirubin complex.
I. The concentration of free Mn^{2+} ions in water solution.
II. The concentration of bilirubin–albumin complex in water solution.

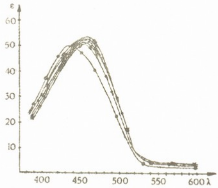

Fig.4. On the pH dependence of absorption spectrum of albumin–Mn^{2+}–bilirubin complex.
▶ ▶ –pH=7.0; × × –pH=7.2;
▷ ▷ –pH=7.4; • • –pH=7.6
▭ ▭ –pH=7.8; ▬ ▬ –pH=8.0
o o –free bilirubin.

We made solution of bilirubin 0.1N NaOH and quickly diminished pH up to pH=6.8. The maximum of absorption of water solution of bilirubin λ=434nm (Fig.2). All the operation took place not in the day light in order to avoid the decay of bilirubin, which could be caused by daylight. We measured the concentration of free bilirubin with the help of the standard curve by the method of spectrophotometry. We took the coefficient of extinction $\varepsilon_{438}=5.2\cdot10^4$ $mole^{-1}cm^{-1}$. We measured free Mn^{2+} concentration in solution according to spectrum EPR with the help of the standard curve.

Bilirubin has two constants of association with albumin. One strong $K_1=7\cdot10^{-9}m^{-1}$ and another comparatively weak $K_2=2\cdot10^{-6}m^{-1}$ [1]. By adding the serum albumin to the water solution we get a complex the maximum of absorption of which is λ=460nm. Fig.2 presents the absorption curves of the human serum albumin with bilirubin complex with different meanings of pH. The absorption curves present in water solution the superposition of the curves of free bilirubin and of bilirubin–albumin complex. As we increase the value of pH the part of absorption curve increases and when pH=8.0 the curve in fact represents the absorption curve of bilirubin–albumin complex.

The serum albumin to Mn^{2+} ions has two centres of binding with the constants of association $K_1=0.63\cdot10^{-4}m^{-1}$; $K_2=0.4\cdot10^{3}m^{-1}$. Besides, the constant of association very much depends on the pH of the solution [2]. On the other side, after our measurements the meaning of the constant of association Mn^{2+} ions to bilirubin does not depend on

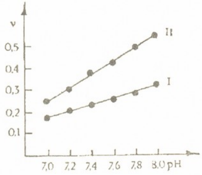

Fig.5 Plots of ν VS pH for HSA.

I. The complex of bilirubin with albumin $\nu = \frac{[Br]_B}{[P]}$

II. The complex of bilirubin with albumin and Mn^{2+} ions $\nu = \frac{[Br + Mn^{2+}]_B}{[P]}$

$[Br]_B$ –concentration of Br bound to protein

[P] –total HSA concentration–$3.6 \cdot 10^{-5}$ M

[Br] –total concentration of bilirubin –$1.8 \cdot 10^{-5}$ M

[Mn^{2+}]– total concentration of Mn^{2+} –$1.8 \cdot 10^{-5}$ M

pH of the solution (Fig.3). By adding albumin to the water solution of bilirubin–Mn^{2+} complex we get the triple complex of bilirubin with albumin and Mn^{2+} ions Fig.4 presents the absorption curves of complex of the numan serum albumin with bilirubin and Mn^{2+} ions with different meanings of pH. As we increase pH of solution the concentration of complex increases and all bilirubin is in the concentration of complex when pH=8.0. Differently of previous case the formation of complex of bilirubin with albumin in the presence of Mn^{2+} ions begins at the acid values of pH and increases in the region of pH characteristic for blood (pH=7.4) (Fig.5).

So we can suppose, that Mn^{2+} ions take part in the formation of albumin–bilirubin complex and it is possible that triple complex is formated through Mn^{2+} bridge mechanism, the existence of which was declared by Marx [1].

In case of colestasis which is caused by bilirubin and, in particular, in case of the acute viral hepatitis and chronic aggressive hepatitis the concentration of Mn^{2+} ions in the blood increases with increasing the concentration of free bilirubin. This fact indicates the existence of albumin – Mn^{2+} – bilirubin complex. On the other side it is known, that the joint function of Mn^{2+} ions and free bilirubin causes the blocking of bile duct and colestasis in the animals.

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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BIOCHEMISTRY

T.Buachidze, L.Topuria

Some Physicochemical and Enzymatic Properties of Fungal Fusant Invertase

Presented by Academician G. Kvesitadze, June 21, 1996

ABSTRACT. **The isoelectric point (pl 4.9) of the invertase from fungal fusant and Km value for sucrose (K_m=58 mM) was determined. The enzyme rapidly hydrolyzes sucrose, but showed little activity of raffinose and was inactive on the other tested disaccharides Cu^{2+}, Ag^{+} and Hg^{2+} cations inhibited the enzyme. The predominant amino acid residues of the enzyme protein were Serine, Glutamic acid and Threonine.**

Invertase (β-frutofuranosidase, EC 3.2.1.26) catalysis the hydrolysis of β (2-1) fructosidal bond into β-fructofuranosides. The enzyme occurs in yeasts, micromycetes, bacteria, actinomycetes and plants. Invertase was also isolated from fungal fusant described in our previous paper [1].

The aim of our study is to continue the investigation of the properties of invertase from fungal fusant.

Materials and Methods

Chemicals. Purified invertase preparation was used from our laboratory, the pharmalate and pl marker protein Kit were purchased from Pharmacia Fine Chemical Corp. (Sweden), sucrose, raffinose, lactose, melibiose ("Serva" Germany), cellobiose, trehalose ("Chemapol" Czekhoslovakia). The other reagents are production of "Reakhim" firm (Russia).

Enzyme assay. The reaction mixture for invertase assay was composed of 0.5 ml 6% sucrose, 0.4 ml 0.05 M Na-acetic buffer, pH 4.6 and 0.1 ml enzyme solution. The reaction was carried out at 50^0 C for 15 min. The amount of reducing sugar released was determined by the method of Somogyi [2]. One unit of the enzyme is defined as the amount of enzyme that catalyses the hydrolysis of 1 μmol of sucrose in 1 min. Specific activity is expressed as units per mg of protein.

Analysis of amino acids. The amino acid composition of invertase was analyzed with a KLA-3 amino acid analyzer ("Hitachi", Japan) after hydrolysis of the enzyme with 5.7 N HCl for 24; 48 and 72 h at 105^0 C in a sealed evacuated tube.

Protein concentration was assayed by the method of Lowry et al. [3] with bovine serum albumin as a standard.

Isoelectric point. The isoelectric point of the enzyme was found by density gradient isoelectric focusing as described by Vesterberg [4] with the use of 1% Pharmalyte (carrier ampholyte), pH 3.5-10.

Results and Discussion

Substrate specificity. 1 ml of 0.1 M substrate solution was incubated with 1 ml of the enzyme and 3 ml of 0.05 M acetic buffer, pH 4.6 for 15 min, at 50^0 C. The enzyme activity toward various substrates (sucrose, raffinose, melibiose, maltose, trehalose, lactose, cellobiose) was determined by measuring the increase in reducing sugars in the reaction mixture. The enzyme was active toward sucrose and raffinose (8% of the activity toward sucrose) but not active toward other disaccharides. These results suggest that the invertase from fungal fusant acts as a β-fructofuranosidase (EC 3.2.1.26) and not as a α-glucosidase (EC 3.2.1.20).

Michaelis constant. Fig. 1 shows the Michaelis parameter of the invertase calculated from Lineweaver - Burk double reciprocal plots. The k_m value for sucrose was found to be 58 mM.

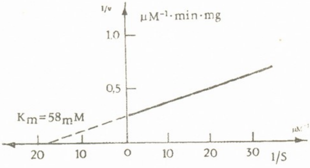

Fig.1

Isoelectric point. The isoelectric point of the enzyme was determined by electrofocusing on an Ampholine polyacrylamide gel plate and equals to 4.9.

Amino acid composition. The amino acid composition of the enzyme is summarized in Table 1. The predominant residues of the enzymic protein were Serine, Glutamic acid and Threonine. The amino acid composition of the invertase from fungal fusant was found to be closely similar to that of the *Aspergillus* awamori enzyme [5].

Effect of various reagents on enzyme activity. The effects of various cations on invertase activity were investigated. The enzyme was preincubated with each reagent in 0.1 M acetate buffer (pH 4.6) at 40^0 C for 1 h and residual activity was assayed. As shown in Table 2, Cu^{2+}, Hg^{2+} and Ag^+ inactivated significantly while other cations and organic compounds inhibited partially. EDTA, DTT and 2 Mercaptoethanol stimulated the activity.

Table 1

Amino acid composition of invertase from fungal fusant

Amino acids	Quantity
Aspartic acid	39
Threonine	54
Serine	65
Glutamic acid	58
Glycine	40
Proline	24
Alanine	36
Valine	22
Cysteine[1]	2
Methionine	10
Isoleucine	19
Leucine	29
Tyrosine	25
Phenylalanine	28
Histidine	23
Lysine	22
Arginine	16
Tryptophan[2]	7
All	519

1 mg of the purified enzyme was used for the analysis of amino acid.

1 Not corrected for destruction.

2 Measured spectrophotometrically.

Table 2

Effect of various cations and chemical compounds on invertase activity

Compound	Concentration M	Residual activity
Cu^{2+}	$5 \cdot 10^{-3}$	0
Mn^{2+}	$5 \cdot 10^{-3}$	30
Hg^{2+}	$5 \cdot 10^{-3}$	0
Ag^{+}	$5 \cdot 10^{-3}$	0
CH_2ICOOH	$2.5 \cdot 10^{-5}$	50
p-CMB	$2.5 \cdot 10^{-3}$	70
EDTA	10^{-4}	122
Dithiothreitol (DTT)	10^{-2}	105
2-Mercaptoethanol	10^{-2}	108
KCN	10^{-3}	90
NaF	10^{-3}	90
NaN_3	10^{-3}	89

Georgian Technical University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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E.Davitashvili, Corr. Member of the Academy N.Aleksidze, G.Alexidze

Lectin-Binding Glycoconjugates and Ca^{2+} -ATPase Activity in White Rat Intestine Cells

Presented June 3, 1996

ABSTRACT. **Galactose-content glycoconjugates dissolved in water-methanol mixture and binding galactoside-specific CS-Gal lectin from *Coriandrum sativum* have been revealed in white rat's intestine epithelium. The protein fraction of intestine cell tritone-extract with MM 35 kD is characterized by higher Ca^{2+}-ATPase activity and CS-Gal lectin binding. Then it inhibits tritone-extract Ca^{2+} -ATPase activity and thus suggests the existence of close link between lectin-binding mechanism and that of rat intestine cell Ca^{2+}-ATPase activity.**

It is known, that some plant lectins have toxicity effect on men and animals after their administration without thermal treatment [1-4]. It was established, that this has been caused by lectin interaction with intestine surface cells preventing absorptive mechanisms and normal functioning of the digestive system [2].

This study revealed the galactose-specific-lectin-binding proteins in rat intestine surface cells. At the same time the influence of CS-Gal lectin on intestine surface cell Ca^{2+}-ATPase activity has been studied. The goal was to study the galactose-specific lectin (CS-Gal) separated from the *Coriandrum sativum*, the vegetables representative, which is often used without preliminary thermal treatment.

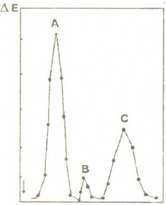

Fig.1. The gel-filtration of white rat's intestine surface cells triton-extract on acrylex P-200 column by 0.1% triton X-100, 20mM Tris-HCl buffer, pH 7.4. Flow rate- 1ml/min.

Lectin was isolated from whole plant Coriandrum sativum by affinity chromatography on tris-acryle-galactose sorbent [5]. White rat's intestine surface cells lectin-binding conjugates were extracted by 0.1% Triton X-100 in 20 mM Tris-HCl (pH 7.4) buffer. The lectin-binding capacity of extracted proteins was studied by hapten-inhibitory technique in hemagglutination conditions [6]. The triton-extract of intestine-surface cells has been fractionated by gel-chromatography on acrylex P-200 column (32cm×0.8) by HPLC ("Knauer") at a flow rate of 1ml per min. Hemagglutination activity of extracts and purified proteins was determined in 2% trypsin-treated rabbit erythrocytes suspension [7] by serial 2-fold dillution of extract in microtiter U-plates. The Ca^{2+}-ATPase activity was estimated according to release of inorganic phosphate quantity in consequence of ATP hydrolise. Ca^{2+}-ATPase activity was calculated as mcM phosphate (1mg protein) 1min. [8,9]. Reaction medium (2.5ml) concluded: Ca^{2+}- (1mM), Mg^{2+}(3mM), ATP (1mM), tissue extract (100mkg) and different concentration of lectin.

Table

The influence of lectin-binding protein fractions (A-1, B-2, C-3) extracted by tritone X-100 from the rat intestine surface cells on CS-Gal lectin hemagglutination activity in trypsin-treated rabbit erythrocytes 2% suspension. A - fraction 1 (initial concentration 20mkg/100mkl) + CS-Gal lectin (4.6 mkg/50mkl)+ Trypsin-treated rabbit erythrocytes 2% suspension (2% Tr.R.E.S.) 50mkl); B- protein fraction 2 (init.conc. 12.8mkg/100mkl) + CS-Gal lectin (4.6mkg/50mkl) + 2% Tr.R.E.S. (50 mkl); C -fraction 3 (init.conc. 17.2mkg/100mkl) + CS-Gal lectin (4.6mkg/50mkl) + 2% Tr.R.E.S. (50mkl); D- intestine surface cells initial tritone-extract (init.conc. 140mkg/100mkl) + CS-Gal lectin (13mkg/50mkl) + 2% Tr.R.E.S. (50mkl); E- CS-Gal lectin (init.conc.9.2mkg/100mkl) + 2% Tr.R.E.S. (50mkl); K- control, hemagglutination buffer + 2% Tr.R.E.S. (50mkl).

A, B, C, D - protein titer, CS-Gal lectin concentration is constant; E- CS-Gal lectin titer."--"- hemagglutination is absent, "+" - hemagglutination, "xx"-lyzis

Experiment variants	Titrators holes number											
	1	2	3	4	5	6	7	8	9	10	11	12
A	x	x	x	-	-	+	+	+	+	+	+	+
B	x	x	-	-	-	-	+	+	+	+	+	+
C	x	-	-	+	+	+	+	+	+	+	+	+
D	-	-	-	-	-	-	-	-	-	+	+	+
E	+	+	+	+	+	-	-	-	-	-	-	-
K	-	-	-	-	-							

By gelfiltration of rat's intestine surface cell triton-extract on acrylex P-200 (Fig.1) three protein fraction with molecular weight 95(A), 35(B) and 16(C) kD has been identified. In control experiments they had no hemagglutination activity in 2% trypsin-treated rabbit erythrocytes suspension. On the other hand they revealed its lectin-binding capacity (table 1) relatively to CS-Gal lectin. In the first hole of microtiter U-plates, where the protein concentration of initial triton-extract or separated protein fractions were high, the hemagglutination activity wasn't revealed. According to inhancing the protein titer, when lectin-binding protein concentration has been decreased, rised free, unbinding lectin concentration and hemagglutination activity was appeared. This effect has been revealed when the relation of lectin to lectin-binding protein concentration was for initial triton-extract 50:1, for the first (A) fraction - 8:1, for the second fraction (B)- 25:1 and for the third (C) fraction -2:1. These results indicate availability of lectin-binding protein in intestine surface cells triton-extract.

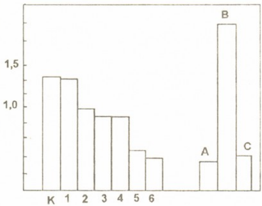

Fig.2. The Ca^{2+}-ATPase activity of white rat's intestine surface cells triton-extract and separate protein fractions (A,B,C) and the influence of CS-Gal lectin on intestine surface cells triton-extract Ca^{2+}-ATPase activity.

Enzymes activity expressed in mcM phosphate /1mg protein/1min. K-control, Ca^{2+}-ATPase activity of initial triton-extract. 1-6 - the influence various concentration of CS-Gal lectin on Ca^{2+}-ATPase activity: 1-0.004mkg/ml, 2-0.4mkg/ml, 3-2.0mkg/ml, 4-4.0mkg/ml, 5-8.0mkg/ml, 6-12mkg/ml. A.B.C. - Ca^{2+}-ATPase activity of lectin-binding fractions

To study chemical character of lectin-binding glycoconjugates in triton-extract methanol-chlorophorm mixture (5:1) was treated by modified method of Patrikeeva-Felch [9]. The glycoconjugates, soluble in water-methanol and chlorophorm didn't induce the hemagglutination of 2% trypsin-treated rabbit erythrocytes suspension, but in presence of CS-Gal lectin (constant concentration) only the water-methanol mixture protein showed lectin-binding capacity.

Considering previous experimental results [12] of the influence of CS-Gal lectin on human erythrocytes shadows Ca^{2+}-ATPase activity and the role of Ca-ions in lectin-receptor binding processes [1] we studied the influence of CS-Gallectin on intestine surface cells triton-extract and on separated protein fractions (A,B,C) Ca^{2+}-ATPase activity. It is established, that CS-Gal lectin (Fig.2) inhibits the Ca^{2+}-ATPase activity. It is necessery to note, that the maximal enzymatic activity has been observed in protein fraction B, which is characterized by higher CS-Gal lectin binding capacity too. The protein fraction A and C do not practically differ.

In conclusion, the CS-Gal lectin binding galactose-content glycoconjugates of white rat's intestine surface cells glycoproteins are characterized by solution in water-methanol mixture and by high Ca^{2+}-ATPase activity. The enzyme activity is strongly inhibited by CS-Gal lectin on lectin's concentration. It is supposed, that inhibition of Ca^{2+}-ATPase activity of rat's intestine surface cells is reflected in digestion processes and in lectins and ingredients transport after their binding with membrane glycoconjugates.

I.Javakhishvili Tbilisi State University

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BIOCHEMISTRY

N.Gabashvili, M.Tsartsidze, A.Gujabidze

On Some Trends in Structural Changes of Various Polymers in the Process of their Damage

Presented by Corr. Member of the Academy D. Jokhadze, February 15, 1996

ABSTRACT. **Some methods of IR–spectroscopic analysis were used to study the structural changes of the polymers in the process of biodegradation of polymeric materials (polyethylene, polytetrafluoroethylene, vinylidenfluoride and hexafluorpropylene copolymer, cellulose, aromatic polyimide, polyethilene-terephthalate and polyvinylchloride) caused by mold fungi. Study of IR–specra of various polymers showed that on one hand the mold fungi depletes the covalent bonds in the existing basing groups (decrease of the respective absorption bands in the IR–spectrum), and on the other – it forms the terminal carboxyl groups the amount of which is increased with the increase of the biodegradation degree. Since this is characteristic of all polymers, the above fact can be used to determine the biodegradation degree for any polymer.**

The polymer structure is changed as a result of methabolites excreted by the mold fungi in the process of their vital activity. The hydroxyl radical which depletes the C–C bonds and forms the oxidized terminal groups was found to play the important role in this [1–3]. From those groups the carboxyl group is of significant importance [4,5]. One of the methods to detect this group is the IR–spectoscopy (with maximal absorption in the range of 1800–1600 cm^{-1}). Such type of damage is presumably characteristic of all polymers containing C–C bond and can be considered in quantitative assessments of the degree of polymer biodegeneration. To check the above assumption the following studies have been carried out.

The polyethylene, polytetrafluoroethylene and hexafluoropropylene copolymers (which do not contain oxygen and do not absorb in the range of 1800–1600 cm^{-1} of IR–spectrum) as well as cellulose (which contains oxygen but does not absorb in the range of 1800–1600 cm^{-1} of IR–spectrum) films where used as an experimental material. Apart from the above listed polymers (such as aromatic polyimide, polyethylene–tetraphthalate and polyvinilchloride) which are characterized by strong absorption ability in the range of 1800–1600 cm^{-1} of IR–spectrum have been studied. The polymers were obtained from the micological area of Research Center of Colchis Climatic Testing (Republic of Georgia, Ozurgeti Region, Village of Shekvetili) where they are tested for resistibility to biodegradationunder the natural conditions (humid subtropics).

Using the method of IR–spectroscopy we investigated the sites with various density of fungi colonies. Predominantly the sites with 5 to 6 points of damage have been investigated [3]. The polymeric films which were obtained from the places other than mycological area served as control. From the polymers under study were prepared the 20 micrometer thick plates of 35x17 mm size. These samples were cleaned from the dirt and micelia of fungi using wet cotton; they were rinsed with distilled water and dried during two days in exiccator on dry silicagel. The IR spectroscopy of the films was conducted using Karl Zeiss (Jena) Firm spectrophotometer "Specord M–80". Before spectrum recording the channels of the spectrometer were subjected to 10 min ventilation with dry air to remove the water vapors. For the samples with strong absorption in the range of 1800–1600 cm^{-1} of IR spectrum the difference spectrum was recorded (i.e. the spectrum which is obtained in one channel of spectrophotometer served as a control, while the spectrum obtained in the other channel – belonged to the plate under study).

The experimental data obtained are provided in Table 1. As it is evident from the Table the polyethylene plate has no absorption in the range of 1700 cm^{-1} of IR spectrum. With the beginning of vital activity of the fungi on the surface of the Film the absorption band appears in the range of 1712 cm^{-1} of IR spectrum (respective maximum of carboxyl C=0 group), its intensity being increased with rise of the damage rate. The above data demonstrate that in the process of biodegradation the terminal C=0 groups appear, conditioned by the consumption of CH_2 groups (the respective maximum for the above group in the range of 720 cm^{-1} is proportionally decreasing).

The trends observed for polyethylene films remain unchanged in case of polytetrafluoroethylene and cellulose (Table 1); at the initial stage of biodegradation absorption maximum in the range of 1716 cm^{-1} appears in polytetrafluoroethylene. At the same time the increase of biodegradation degree is accompanied by decrease of absorption maximum in the ranges of 780, 742 and 720 cm^{-1} which corresponds to the amorphous phase of polytetrafluoroethylene. This fact is indicative of depletion of C=C bonds and formation of terminal carboxyl group in the polymer.

Significant interest is IR spectrum of vinylidenfluoride and hexafluoropropylene copolymers damaged by mold fungi under natural conditions: the structure of the above polymers includes functional groups typical for polyethylene and polytetrafluoroethylene. Here as in case of the previously discussed polymers the maximum absorption in the range of 1708 cm^{-1} is observed at the initial stage of the fungus vital activity on the surface of the polymer. Its intensity is increasing with the increase of biodegradation degree. At the same time the decrease of the maximal absorption intensity in the ranges of 3020 and 2980 cm^{-1} is vinilidenfluoride groups (Table 2). The quantitative decrease of fluorethylene and ethylene groups is also observed being expressed in decrease of maximal absorption intensity in the ranges of 740 and 720 cm^{-1} of IR spectrum.

It should be noted that between the quantitative increase of terminal carboxyl groups and decrease of basic functional groups of the polymer exists directly proportional relationship. One can judge about polymer biodegradation degree by the increase of respective absorption maximal intensity of newly formed groups as well as by the decrease of respective absorption maximal intensity of the basic functional groups. Comparison of the IR spectra of the mold fungus damaged and intact cellulose plates shows the disruption of the basic chain of the cellulose in case of mold fungi and appearance of the respective absorption maximum (in the range of 1718 cm^{-1}). With increase of biodegradation degree the intensity of absorption maximum increases (Table 1).

Table 1

Change in the Intensity of Some Characteristic Bands in the IR Region Mold Fungi in the Process of Polyethylene, Polytetrafluorethylene and Cellulose Biodegradation

NN		Degradation Degree in Points	Intensity of Absorption Bands (D, cm^{-1})				
1	Polyethylene			D_{1712}		D_{720}	
		0		0		2,15	
		3		0,36		1,95	
		5		0,57		1,70	
2	Polytetrafluro–ethylene		D_{1716}	D_{1140}	D_{780}	D_{742}	D_{720}
		0	0	1.418	0.251	0.313	0,355
		3	0.032	2.426	0.240	0.307	0,338
		5	0.124	1.577	0.027	0.261	0,294
3	Cellulose			D_{1718}			
		0		0			
		3		0,653			
		5		0,765			

Table 2

Changes of Some Characteristic Bands in Difference Spectrum of Aromatic Polyimide, Polyethyleneterephthalate and Polyvinylchloride Degradation Caused by Mold Fungi under the Natural Conditions

NN		Degradation Degree in Points	Intensity of Absorption (D, cm^{-1})				
1	Aromatic polyimide			D_{1778}	D_{1645}	D_{1260}	D_{784}
		0		0	0	0	0
		3		–0,108	0,089	–0,052	–0,090
		5		–0,270	0,102	–0,136	–0,121
2	Polyethylene–terephthalate		D_{1712}	D_{1330}	D_{730}		
		0	0	0	0		
		3	0,0123	–0,043	–0,081		
		5	0,187	–0,083	–0,105		
3	Polyvinyl-chloride		D_{1778}	D_{1645}	D_{1260}	D_{784}	
		0	0	0	0	0	
		3	–0,108	0,089	–0,052	–0,090	
		5	–0,270	0,102	–0,136	–0,121	
4	Vinilidenfluoride and hexafluoropropylene copolymers		D_{3020}	D_{2980}	D_{1708}	D_{740}	D_{721}
		0	1,244	0,996	0	2,21	1,89
		3	1,203	0,871	0,075	1,98	1,76
		5	1,087	0,818	0,241	1,44	1,62

It is important to study the polymers which being sufficiently resistant to the aggressive media are at the same time easily degraded by the mold fungi. To those belong polyethyleneterephthalate and aromatic polyimide. The method different from the above described was used to study their IR spectra in the region of 1800-1600 cm^{-1}. These polymers and polyvinylchloride have a pronounced absorption band in the region of 1800–1600 cm^{-1}. For this reason the control film was placed in one channel of the spectrophotometer and IR spectra of intact films were recorded at the lowest scanning speed. The obtained difference spectra were numbered and entered into the computer. During experiment the IR spectrum was recorded for the control plate, the plates with various degree of damage caused by the mold fungi and the difference between the spectra difference of the polymers before and after their biodegradation was calculated using the special software of the built–in computer of the spectrophotometer.

The obtained results using the above described method are given in Table 2. From the Table it is evident that in case of polyethyleneterephthalate the intensity of the respective absorption maximum of the terminal carboxyl group (1712 cm^{-1} region) increases with the decrease of the intensities of the respective absorption maximums of CH_2 groups (1330 and 730 cm^{-1}). The similar results were obtained in case of the biodegradation of aromatic polyimide and polyvinylchloride films (Table 2).

Thus, the analysis of the IR spectra in the process of biodegradation of the above polymers demonstrated the following common trends: on one hand the mold fungi deplete the covalent bonds in the existing basing groups (decrease of the respective absorption bands in the IR spectrum), and on the other – from the terminal carboxyl groups the amount of which is increased with the increase of biodegradation degree. Since this latter is characteristic of all polymers, the above fact can be used to determine the degree of biodegradation for any polymer.

I. Javakhishvili Tbilisi State University

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BIOCHEMISTRY

R.Goguadze, N.Mikiashvili, M.Chipashvili, Corr.Member of the Academy N.Aleksidze

Comparative Analysis of Biogenic Amines Distribution in White Rat Brain at Different Forms of Aggression

Presented February 15, 1996

ABSTRACT. **The distribution of biogenic amines in different parts of white rat brain after pilocarpine and spontaneous aggression have been investigated.**

The mouse-killing rat formation has been demonstrated on the ground of pilocarpine-induced aggression (12.5, 25mg/kg/day).

Pilocarpine-induced mouse-killing activity of rats appeared to be a form of behavioral pathology that differed from the species - specific predatory response. Significant differences have been discovered in the biogenic amines distribution at the different forms of pharmacological and spontaneous aggressions.

At present various psychotropic drugs are extensively used for the regulation of emotional state and for the reabilitation of people that have been under the emotional stress pressure.

That leads to more detailed investigations of various drug actions on emotional reactivity and on psychophysiological behaviour.

Proceeding from the above-mentioned biochemical evaluation by pharmacological drug pilocaprine induced aggression is of particular importance. Several model systems have been proposed for aggressive behavioral investigations. The most suitable seems to be the model of "killer" rats. It is displayed in definite emotional reactions of the animal on extremal stress-factors and pharmacological agents [1].

In the previous paper active participation of olfactory balb in such processes as the sharp strengthening of the aggressive behaviour has been shown. Analogical reactions were developed at the pilocarpine and phenamine administrations to the animals [2,3,4]. Further more, it has been demonstrated that the major role in the development of the aggressive behaviour belongs both to serotoninergic and noradrenergic systems. However their participation in the development of the aggressive state regarding the defined nuclei of the lymbic system is not still clear.

Thus, this paper deals with comparative studies of quantitative distribution of biogenic amines in the different parts of brain at spontaneous and pilocarpine-induced aggressions. Together with the development of aggressive state, the distribution of biogenic amines have been evaluated in different parts of nonlinear white rats brain.

For all the experiments male white rats, weighing 120-200g were used. All the animals were divided into four groups:

I group - animals were injected with pilocarpine (PC) intraperitoneally daily for 21 days with 12.5 mg/kg concentration .

II group - doubled dose PC (25 mg/kg) was administered during 21 days.

III group - Physiological saline (1ml) was injected for 21 days; animals were periodically irritated mechanically.

IV group - control one, administered only with physiological saline (1 ml).

Before drug administrations all rats were tested for mouse-killing according to the method of Plotnik et al [5]. The test was repeated 60 min after the injection (once per 7 days). Sometimes the "killer" rat was obtained after the first pilocarpine administration, but that was not statistically confirmed.

Biogenic amines were determined by high pressure liquid chromatography. The system was applied by electrochemical detector [6]. Protein concentrations were determined by the method of Lowry et al [7].

The results are illustated in Figures 1 and 2. Figure 1 shows that experimental rats develope aggressive behaviour and defensive-motor activity after 1.5 hour of the first pilocarpine administration.

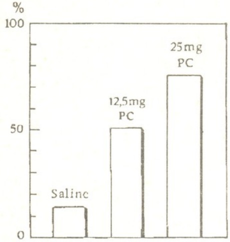

Fig.1. Dependence of mice "killer" rats induction from PC concentration

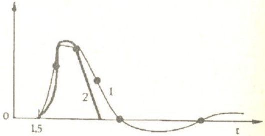

Fig.2. Dynamic of different forms of aggressive state development
1 - Pilocarpine induced aggression
2 - Spontaneous aggression

Drug chronic injections induced the mouse-killing rats (from 50% to 75%) in a close-dependent proportion. Inhancement of PC concentration (25mg/kg) more effectively induced mouse-killing rats approximatly to 75%. However PC administration for 21 days gradually decreased locomoter activity and rearing activity of rats. As the PC injections at the concentration of 25 mg/kg induced sharp depression in rats. It was decided to use the drug of 12.5 mg/kg concentration for comparative biochemical studies of biogenic amines distribution. It should be pointed out, that the topography of the killing response displaying by PC-treated animals differed clearly from that seen in natural mouse-killing rat [2,8]. Due to importance catecholamines in time dependent development of aggressive state in animals at critical periods (before, after 1.5 hour and 24 hour intraperitoneally drug administrations), we have investigated the quantitative distribution of dophamine (DA), noradrenalin (NA) and serotonin in brain regions; these are responsible (midbrain, hypothalamus, hippocampus) for development of serotonin - dependent aggression. The results are presented in Table 1.

The Biogenic amines content in different parts of the rat brain
($\mu g/g$ tissue)
(M±m)

Brain structure	Biogenic amines	Control	I Group [after final injection PC]	III Group
midbrain	Dophamine	0.32±0.04*	0.9±0.03**	0.8±0.3**
	Noradrenalin	0.38±0.02*	1.2±0.18*	0.68±0.22
	Serotonin	0.80±0.05**	0.4±0.02*	0.50±0.06**
hypothalamus	Dophamine	0.58±0.15*	0.73±0.04**	0.95±0.07**
	Noradrenalin	1.85±0.22**	2.5±0.20*	2.2±0.13*
	Serotonin	1.12±0.37	0.45±0.06*	0.6±0.02**
hippokampus	Dophamine	0.40±0.12	0.6±0.04*	0.45±0.02*
	Noradrenalin	0.85±0.06**	1.00±0.12*	1.10±0.17*
	Serotonin	0.60±0.14*	0.3±0.06*	0.20±0.01*

* $p < 0.05$
** $p < 0.01$

The major cell body clusters of serotonin neurons are mostly located in midbrain, whereas their acsonal parts and synaptic termini are found in hippocampus and hypothalamus [9]. The investigated regions of brain in intact animals have the following sequence regarding to NA concentrations: hypothalamus> hippocampus>midbrain. The increased level of NA and DA in different parts of brain at the development of aggressive state and the decline of their concentrations in parallel of normalization of mottor activity is considered as major regularities of their quantitative distributions. It have been established that NA quantity exceeds DA level. This result is supported by current data [1]. In reply on PC-administration (1.5 hour) the quantity of NA sharply increases in all three regions of brain. At the spontaneous the amount of NA increases mainly in hypothalamus and hippocampus. The data on midbrain are not statistically significant.

It should be noted, that there are no inner sources of catecholamines in the hippocampus, but they penetrate from the outward.

Both NA and DA evoked inhibitory effect on the same hippocampal cells [10]. Later on we have examined two phases in development of aggression:

1 - the first beginning phase has an adaptive role;

2 - the second was of a pathogenetic character.

DA level augmented at PC administration. In contrast to NA the increased DA concentration was determined in midbrain during 1.5 hour after the PC injection. In case of spontaneous aggression increase was detected in the hypothalamus and midbrain, but it was remained on normal level in the hippocampus. It seems that balanced functional activity of pre- and postsynaptic receptors, characterizing normal cases, were disrupted and led to changes in motor activity. It was also established that the serotonin concentration decreased both at PC-induced and spontaneous agressions. In midbrain (I group) serotonin concentrations changed from 0.80 μg/g, to 0.40 μg/g that made up 50%. In the same structure of intact animals brain serotonin quantity decreased on 37,5% (Table 1).

It must be noted, that during activity of serotonin metabolism ferments and consequently serotonin level differ in various structures of rat brain and it correlates with the differential distribution of serotonin system. These results suggest that serotonin depletion should facilitate nonspecific killing reactions, but it is not sufficient to induce specific predatory behaviour in rats [8]. Absence of mouse-killing in the third experimental group indicates that alteration in biogenic amines distribution is not the only cause for the induction of mouse-killing, although it is possible contributing factor. The topography of the killing response was displaying in PC treated animals differed clearly from that seen in natural mouse-killer rats. Thus, we should argue that although serotonin may be involved in muricide, its involvement is probably indirect.

Basing on the present data it should be concluded, that the development of animal aggressions is connected with the disruption of biogenic amines distribution in brain, with the activity of monoaminoxidase and physiological mechanisms, involved in the formation of aggressive conduct.

I.Javakhishvili Tbilisi State University

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ZOOLOGY

E.Kvavadze

Shape of Genital Setae-A Taxonomic Character of the Genus Eisenia (*Oligochaeta: Lumbricidae*)

Presented by Corr.Member of the Academy B.Kurashvili, June 21,1996

ABSTRACT. **By the scanning electron microscopy genital and locomotor setae of 25 species of earthworms have been studied. It is established that species of the Eisenia genus are equipped with a trihedral genital setae and the sculpture of locomotor setae is quite definite and different for various species.**

In spite of the effort of many generations of zoologists, the system of the *Lumbricidae* family is far from being perfect. The majority of criteria used for taxa separation have almost exhausted their possibilities. That is why there arose the necessity of searching new taxonomic characters. This paper is devoted to the study results of the structure and sculpture of locomotor and genital setae of the earthworms belonging to the genus *Eisenia*.

As it is known, there are two types of setae in earthworms: locomotor and genital [1,2]. On some segments of the body front part, especially in the region of clitellum, there are genital setae. In the *Lumbricidae* family, unlike other *Oligochaeta*, the genital setae are not related to the genital pores topographically. The genital setae are longer, some of them are pointed in the distal part with longitudinal furrows. They are usually located on glandular papillas. The data on structure and sculpture of the genital setae of the *Lumbricidae* family are presented in Stephenson's monograph [2] mentioning that genital setae of *L.terrestris* and probably, those of all the species of the *Lumbricus* genus are very much alike to the locomotor ones. However unlike the latter, they are longer and there are broadenings at the end of the distal part resembling a slight link near the tip. According to Michaelsen two types of the genital setae are typical for *Dendrobaena mariupoliensis*. Some of the setae are strikingly similar to the locomotor setae both in shape and sculpture. But unlike the latter, they are longer and their sculpture is sharper, more regular and deeper. Other genital setae are thin, pointed with four ridges in their distal part. That is why Stephenson considers that tetrahedral genital setae are characteristic for most of the species of the *Lumbricidae* family. Tetrahedral genital setae are also mentioned for the *Eiseniella* (*E.tetrahedra*) genus [3].

It should be noted that majority of further generation researchers engaged in the systematics of the *Lumbricidae* family paid too little attention to the structure and sculpture of the earthworms setae. S.M.Svetlov's work is an exception [4]. His *E.magnifica* has S-shaped locomotor setae, uneven at the distal end and with a nodule in the distal fourth. The genital setae are more straight, deprived of a nodule, with pointed end and trihedral shape of the cross-section in the distal third.

The observed peculiarities of the structure of the genital setae made us study the structure of the setae in detail using up-to-date research methods. Structure and sculpture of locomotor and genital setae of 25 species of 2 genera of the *Lumbricidae* family were studied under scanning electron microscops (SEM). 20 endemic and 3 widely spread species of the *Eisenia* genus were studied which cover the main part of the genus from Japanese Lakes to the Central Europe.

Special interest should be paid to the fact that out of 23 species of the *Eisenia* genus, four species have tetrahedral genital setae, while the rest ones have trihedral setae (Table). As was mentioned above, the trihedral genital setae are also typical for E.magnifica [4]. The analysis of the data from literature together with the results of the electron microscopic studies have shown that 20 species of the *Eisenia* genus are supplied with trihedral genital setae (Fig. B, D, J) and hence one can claim that trihedral shape of the genital setae has a stable character.

Table

Shape of the genital setae of earthworms from the *Lumbricidae* family

№	Earthworms	Shape of the genital setae	
		trihedral	tetrahedral
1.	*E.andrei Bouche*, 1972	+	
2.	*E.atlaviniteae Perel, Graph.*, 1985	+	
3.	*E.brevicliteluta Kvavadze, 1985*	+	
4.	*E.colchidica Perel*, 1967	+	
5.	*E.djurganica (Perel*, 1969)	+	
6.	*E.foetida* (*Savigny*, 1826)	+	
7.	*E.grandis grandis Mishaelsen*, 1907		+
8.	*E.grandis ganjiensis Kvavadze*, 1985		+
9.	*E.hydrophilica Kvavadze*, 1973		+
10.	*E.intermedia (Michaelsen*, 1901*)*	+	
11.	*E.iverica (Kvavadze*, 1973*)*	+	
12.	*E.japonica (Michaelsen*, 1891*)*	+	
13.	*E.lagodechiensis Michaelsen*, 1910	+	
14.	*E.malevici Perel*, 1962	+	
15.	*E.nordenskioldi (Eisen*, 1879*)*	+	
16.	*E.rosea (Savigny*, 1826*)*	+	
17.	*E.thanarae Kvavadze*, 1983		+
18.	*E.trascaucasica Perel*, 1967	+	
19.	*E.sibirica Perel*, 1985	+	
20.	*E.spelea (Rosa*, 1901*)*	+	
21.	*E.uralensis Malevic*, 1950	+	
22.	*E.ventripapilata Perel*, 1985	+	
23.	*E.lucens (Waga*, 1857*)*	+	
24.	*A.caliginosa caliginosa (Savigny*, 1926*)*		+
25.	*A.brunocephala Kvavadze*, 1985		+

Therefore, the diagnosis of the genus can be formulated in a new way. It's main characters are trihedral genital setae, very closely paired setae and location of spermatothecae pores on the dorsal side near the medial line. Due to introduction of the new taxonomic character, the genera *Eisenia* and *Allolobophora* have become more isolated. The border between these genera, which had vanished in Pop's system [5] reappeared because the results of these investigations showed that *A.caliginosa* caliginosa and *A.brunocephala* have tetrahedral genital setae rather than trihedral (Fig. C, E, I). Combination of this character with very closely paired setae as well as allocation of the pore of the spermatothecae on the line of the cd setae enables to

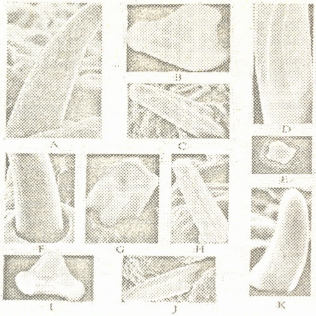

Fig.1. A-locomotor seta of *E.nordenskioldi* 2^c x 500;
B-genital seta of *E.rosea* 27^a x 2500;
C-genital seta of *A.brunocephala* 16^a x 2500;
D-genital seta of *E.nordenskioldi* 16^b x 2500;
E-genital seta of *A.caliginosa caliginosa* 11^a x 2500;
F-locomotor seta of *E.sibirica* 22^a x 2500;
G-genital seta of *E.grandis grandis* 25^a x 2500;
H-genital seta of *A.caliginosa caliginosa* 11^b x 2500;
I-genital seta of *E.intermedia* b on the clitellum x5000;
J-genital seta of *E.intermedia* a on the clitellum x2500 (Zhiguli Reservation);
K-locomotor setae *D.ilievae* 12^b x 2500.

identify clearly the *Allolobophora* genus. It should be emphasized that our factual material showed the presence of trihedral genital setae in the earthworm *E.rosea* (Savigny, 1826) transferred by Pop [5] from the *Eisenia* genus to the *Allolobophora* because of lack of purple pigmentation. In Gate's system [6] there was no place for *E.rosea* in this genus either. Perel [7] assigned *E.rosea* to the Nicodrilus genus, since *E.rosea* has U-shaped nephridial bladders and Morren's glands on the 10th segment. However Omodeo and Rota [8] do not agree with this viewpoint and again assign *E.rosea* to the *Allolobophora* genus. Among present day scientists only Reynolds [9] considers *E.rosea* within the *Eisenia* genus. Not only the shape of the genital setae, but also biochemical data show that *E.rosea* has nothing common with other species of the *Nicodrilus* genus, in particular by the disk electrophoresis method were investigated. The comparison of the similarity indices of the common proteins of the *N.roseus* and *N.longus* species has demonstrated a similarity level corresponding to the species of different genera. It was concluded that *E.rosea* does not belong to the *Nicodrilus* genus [10].

Therefore we can claim that *E.rosea* is genetically related to the species of the *Eisenia* genus and it should be considered as a member of this genus, while the *Nicodrilus* genus should, probably be cancelled. The rest species of the *Nicodrilus* genus should be transferred to the genus *Allolobophora*. Tetrahedral genital setae are also of earthworms from the genus *Dendrobaena* [11]. The species *E.grandis*, *E.grandis ganjiensis*, *E.thamarae*, *E.hydrophilica* having tetrahedral genital setae (Fig. G) and closely paired setae and spermatothecae pores on dorsal side (near the dorsal pore line) should be excluded from the *Eisenia* genus and assigned to the genus *Dendrobaena* or they may probably constitute a separate genus.

Institute of Zoology
Georgian Academy of Sciences

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EXPERIMENTAL MEDICINE

T.Khomasuridze, J.Kristesashvili

The Routes of Genital Infections Distribution in Girls

Presented by Corr. Member of the Academy T.Decanosidze, May 27, 1996

ABSTRACT. **The goal of present investigation was finding the routes of genital infectious distribution in sexually inactive (virgine) girls and adolescents.**

The study involved 116 female patients aged 2-18 with vulvovaginitis. Their mothers were examined also.

The results of bacteriological investigation showed that etiologic factor of genital infection was identical in 81.2 % (age group I), 53.6 % (age group II) and 56.2% (age group III).

Many studies are being devoted to finding the routes of genital infection distribution recently.

The major issue which is still debatable is the distribution of genital infection in sexually inactive girls. The data obtained from various studies to solve this problem are rather controversial.

Of interest is the fact that by the present International Classification the genital infections are defined as Sexually Transmitted Diseases (STD).

If this view is considered right, then the existence and nosology of genital infections in sexually inactive patients (girls), the incidence of which among the gynecological diseases of girls is as high as 40-60%, should have been practically excluded which gives the whole problem the tint of casuistry.

Great majority of the scientists [1,2] believes that of all possible routes of genital infection distribution (sexual, social, hematogenic, etc.) the sexual route plays the most important role.

A number of authors [3,4,5] believe, that distribution of genital infections occurs only via sexual route. Also there are data on transmission of the genital infections by vertical route (during the labor) [6,7,8]. However some researchers [9,10] pay significant attention to the social route of genital infections distribution and believe that this route deserves a deeper and more consistent research.

The versatility of the results of the studies performed to our mind is caused mainly by the selection of the material to be investigated: only few studies involved sexually inactive girls, but even in these cases the source of genital infections was not determined.

One of the most important goals of our investigation was to findg the routes of genital infections distribution in sexually inactive patients referring to their age (and hence their anatomic and physiological peculiarities), the characteristics of the causing agent of the infection and existing cultural and social conditions in complex rather than investigation of the route of genital infection distribution as such.

While examining a patient special attention was paid to the examination of external genital organs to exclude the girls with the history of sexual intercourse or raping.

Examination of the patients' mothers was considered to be mandatory to ensure the right direction of the investigation. When compiling anamnesis special attention was paid to the complaints of both patients and their mothers; duration, of the complaints periodicity and acuteness. The parallels between the onset of the first symptoms of the vulvovaginitis and inflammatory diseases of mother (both current and past) were determined. The everyday life and social conditions of the patients were thoroughly studied.

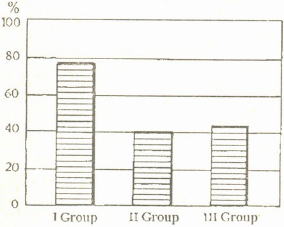
The parents were questioned about their family conditions (the way of washing, storing and using linen, underwear, towels, sponges). It is noteworthy that on interviewing 108 patients' mothers it was found out that in 90 cases the rules of hygiene were not observed.

Laboratory studies included bacterioscopic and bacteriologic examination of vaginal smears obtained from both the patients and their mothers.

The results of the examination of the patients and their mothers were compared and in case of necessity they were administered simultaneous therapy.

In group I which included patients within age group of 2 to 8 years 16 girls and 16 mothers were studied .

Mothers of 14 patients had symptoms indicative of genital infection (leucorrhagia, itch, dysuria phenomenon). Ten women were complaining of these symptoms periodically. Irrespective of the therapy they were receiving the duration of these complaints was from 3 months to 3 years (chronic course). Four women had the symptoms of genital infection within the recent 3 months. Two women did not have complaints indicative of genital infection.



In the above group bacterioscopic and bacteriological examination of the vaginal smears of the patients and their mothers demonstrated that in 13 cases etiologic factor proved to be identical (in 6 cases Trichomonas was a leading etiologic factor, in 4 cases - fungi, in 3 cases - pathological microbes, in 1 case - Chlamydia) (Fig. 1).

As for group II, it consisted of 28 girls (of prepubertal age) over 8 years but without established menstrual function. Their mothers (28 women) were also examined. Of 28 examined women 21 had complains characteristic of genital infection. Eight women had the history of repeated periods of these complaints for more than 3 months (chronic course), while in 6 cases the duration of these symptoms did not exceed 3 months. Seven women did not have any complaints characteristic of genital infection.

Analysis of the vaginal smear of the patients and their mothers from the above group demonstrated that in 15 cases etiologic factor was identical. (In 7 cases it was fungi, in 5 - pathogenic bacteria, in 3 - Trichomonas).

In group III 72 patients of pubertal age were examined (up to 18 years) with established menstrual function, and their mothers (64 women).

Of 64 women to be examined 45 when interviewed mentioned the symptoms characteristic of genital infection. In 35 cases infection had chronic nature, in 1st case it had the acute form. In 19 women of the above group symptoms indicative of genital infection were not found.

Analysis of the vaginal smear of the patients and their mothers from the above group demonstrated that in 36 cases etiologic factor was identical. (In 15 cases the etiologic factor was Trichomonas, in 12 - pathogenic bacteria, in 8 - .fungi and in 1 - Chlamydia).

Expressed in percents these data testify that in group I examination of the patients and their parents demonstrated that etiologic factor of genital infection was identical in 81.2% of cases, the same parameter for group II was 53.6% and for Group III - 56.2%.

The above data provide a sound ground to suppose, that in those patients the causing agent of infection was transmitted via social route rather than sexual one (Fig. 1).

Hence, comparison of the data obtained as a result of examination of the patients and their mothers showed that in group I common etiologic factor was found in 81.25% of cases. This parameter confirms our opinion about the possibility to transmit genital infections via social route, since at this stage of sexual maturity (neutral period) there are significant factors facilitating the penetration of infectious agent into genitals of the girl. First of all this is related to the everyday life conditions: at this age girls are in a closer contact with their mothers, they often sleep in their beds. Important role is played by the anatomic and physiological peculiarities of the girls which involve the absence of pubic hair, flat pudental lips, alkaline medium of the vagina, etc. The childhood infections which change immune status of the organism also play a provocative role.

The patients of groups II and III belonged to the cohort of school children. In these groups the identical etiologic flora expressing index was found in 53.6% and 56.25% of cases, respectively. Irrespective of the fact, that these patients are socially more active, if they do not preserve the rules of hygiene, they do get infected, though the chances to catch infection by this route are lower compared to group I.

As our studies have demonstrated, the results in all three groups are rather significant and their analysis considering age-related anatomic and physiological peculiarities of the girls once more confirms the possibility of genital infections transmitting via social routes in sexually inactive girls. Even more, basing on the results obtained it can be claimed that in sexually inactive girls the major route of genital infections distribution is social.

I.Zhordania Research Institute
of Human Reproduction

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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EXPERIMENTAL MEDICINE

Sh.Gogishvili, G.Lursmanashvili, K.Vacharadze

Influence of He-Ne Laser on Cytological Changes in Bronchial Stump After Pulmonectomy in Patients With Fibrocavitary Tuberculosis

Presented by Academician T.Dekanosidze, October 22, 1996

ABSTRACT. **Correlation has been determined between recovering process in bronchial stump (BS) after operation and cytological picture of bronchoalveolar lavage (BAL) specimens. Parameters have been found out that can be used for early diagnostic of postoperation complications. Effectiveness of Low Energetic Infrared Laser has been confirmed during scar formation in bronchial stump. Guideline has been worked out of insertion in treatment of this procedure and clinical investigations that will significantly decrease frequency of such dangerous complications as bronchopleural fistula (BPF).**

One of the most dangerous complications among postoperation complications of patients with Fibro-Cavitary Tuberculosis (FCT) after pulmonectomy is BS insufficiency followed by BPF formation. Some 15-25% patients have this complication [2;5;8] and mortality indicated in 40-50% [4;6;9].

Reasons of BS insufficiency after pulmonectomy are divided in four main groups [3;5;7]:

1. BS necrosis due to abundant scar formation or by ischemia induced by bronchial wall squeesing with suture.
2. Endobronchitis in resection site, destructive changes of bronchial wall.
3. Infection of pleural cavity, not complete isolation of BS.
4. Immunocompromised patients.

The aim of our study was to observe recovering process in BS after pulmonectomy by cytological investigation of specimens taken from bronchial lumen; also to determine effectiveness of endobronchial He-Ne Laser beam in postoperation period [1].

128 patients with FCT located in Tbilisi G.Mestiashvili's Memorial Phthisio-Pulmonologic Hospital whom had been done pulmonectomy were investigated in 1985-95.

All patients were undergone to flexible fiberoptic bronchoscopy (FFB) with "Olimpus" (Japan) bronchoscope. Also tissue specimens from bronchial lumen for cytological investigations by BAL or biopsy were obtained.

Damage of bronchia was found in 118 (92.2%) patients, among them in 16 (12.5%) it was of specific character, in 80 (62.5%)-non-specific, and in 22(18%) the reason of damage was mixed.

39 (30.5%) patients had 1st degree endobronchitis, 51 (40%) - 2nd degree and 28 (21.9%) - 3rd degree by Lemuan's classification. Damage of bronchi was total in 78 (66.1%) patients and local in 40.

Course of disease was acute in 92 (77%) patients and chronic in 26 (33%).

Patients were randomly divided in two groups: 1st group included patients undergone FFB with transbronchial drainage (TBD) and 2nd group - undergone FFB with TBD plus endobronchial laserotherapy (EBLT). FFB was done in both groups on 3rd, 7th, 14th, 25th and 40th days after operation. 0.1ml of BAL material was transferred on specimen glass, dried, fixated by Leishman's solution and stained by Romanovski's method. Cells were counted by Microscope "LOMO-2" with 90-fold magnification. Couting was done in three visual areas, 100 cells were counted in each and then the average was calculated. Statistical analysis of data was done. Normal bronchial cytology was determined in 10 healthy volunteers (Control group).

He-Ne infrared laser device "Alok-1" (USSR) with power - 1.6 mlv-w. and wavelength - 0.63 mkm. was used for EBLT. Laser beam was led from device by flexible monofibrous beam-conductor (0.5 sq.mm.) that was freely conducted in FFB "Olimpus" drainage channel. Bundle of beam was dispersed on 0.5 sq.cm.area in 1.0 cm. length by means of lens located on the end of beam-conductor. Duration of each procedure was under 3-5 min.

By the end of each procedure radiation dose was calculated by equation: $E=NT(1-P)/S$ where:

E - single dose of radiation (Joule/sq.cm.);

N - power of laser beam on the end of beam-conductor;

T - exposure (sec.);

S - surface of radiation area;

P - constant of radiation absorption of exposed tissue;

Cytological investigation of site of bronchus resection has shown that sutured stump can be hermetic to 10-14 days, then laceration of tissue around the clips takes place, but at that time granilation tissue followed by scar that prevents wound dehiscence develops. Hematoma, edema and neutrophil infiltration of stump area is indicated on 1st - 3rd day after operation. Metaplasia of columnar epithelium to stratified squamous epithelium takes from 3rd day; at that time also mitotic acticity of cells of basal layer is indicated. Damaged epithelium almost completely recovers, exudation decreases and content of fibroblasts and collagen increases in suture sites on 7th day after operation. Stratified epithelial layer completely develops on 14th day after operation, but there are some ischemic areas with dystrophic changes, which without exacerbation heal and mature scar is formed in stump area on 30th day after operation. In case of exacerbation during 2-4 weeks micronecrotic regions develop that lead to bronchopleural complications.

Performed investigations have shown positive correlation between data obtained by BAL and severity of damage in bronchial lumen: more severe is inflammatory-destructive process, more severe are disturbances in cytogram. Particularly in cases of 2nd - 3rd degree endobronchitis the number of alveolar macrophages and lymphocytes decreases (40% fewer than normal number, $p<0.001$; 80% fewer than normal number, $p<0.01$; respectively), but the number of neutrophils and degenerative epithelium increases (twice as many as normal, $p<0.001$; 2.7 times as many as normal, $p<0.001$; respectively). Difference in neutrophil infiltration in postoperation patients with BPF from normal is 25%($p<0.05$) on 5th-8th day after operation and achieves 100%($p<0.001$) on 15th day. At that period also increases the number of degenerative epithelium (16% instead of 7%, $p<0.01$). Significant increase in number of fibroblasts and lymphocytes ($p<0.001$) in BAL cytogram on 8th day after operation indicates activation of stump repairing process and significantly decreases the risk of stump insufficiency (Table 1).

Table

Cytological picture of BAL in BS while recovering

	Cytogram	Stump heading without complication	Development of BPF in stump area	P
N	Number of observations	39	6	
3rd day	Alveolar macrophage	62.15±3.34	57.33±2.17	p1-2<..05
	Neutrophil	20.28±1.18	24.78±1.14	p1-2<. 01
	Epithelial cell	14.72±1.07	16.35±2.05	p1-2<. 05
	Lymphocyte	2.85±0.33	2.57±0.31	p1-2<. 05
	Fibroblast	-	-	-
N	Number of observations	39	5	
8th day	Alveolar macrophage	65.11±2.17	60.56±2.38	p1-2<.05
	Neutrophil	16.36±1.04	25.18±1.72	p1-2<.001
	Epithelial cell	10.23±1.22	10.07±1.13	p1-2>.1
	Lymphocyte	3.25±0.39	3.18±0.12	p1-2<.05
	Fibroblast	5.05±0.22	1.01±0.11	p1-2<.001
N	Number of observations	36	3	
15th day	Alveolar macrophage	72.17±2.78	56.24±1.99	p1-2<. 01
	Neutrophil	10.27±1.37	23.60±1.01	p1-2<.001
	Epithelial cell	8.33±1.05	17.76±1.38	p1-2<.01
	Lymphocyte	6.12±0.78	2.40±0.57	p1-2<.001
	Fibroblast	3.11±0.54	-	-
N	Number of observations	35	2	
25th day	Alveolar macrophage	75.24±3.17	55.01±1.94	p1-2<.001
	Neutrophil	5.34±1.16	25.39±1.23	p1-2<.001
	Epithelial cell	7.11±1.19	16.30±1.17	p1-2<.01
	Lymphocyte	6.14±0.98	2.25±0.37	p1-2<.001
	Fibroblast	6.17±1.38	1.05±0.13	p1-2<.001

EBLT effectiveness was confirmed according to performed investigation (Table 2). EBLT on 5th-8th day after opertion decreases neutrophil infiltration in 2nd group 1.8 times as much as in 1st group ($p<0.001$). In 2nd group fibroblasts appear on 8th -10th day after operation and the number of lymphocytes increases 2-3 times as many as in 1st group ($p<0.01$). Concentration of alveolar macrophages, lymphocytes and fibroblasts increases in stump area on 15th- 20th day after operation that indicated active proliferation in BS. Finally formation of mature scar takes place. 25-30 days were necessary for stump healing in 2nd group patients with EBLT while 35-45 days in 1st group.

Table 2

EBLT influence on process in main BS (cytological investigation of BAL) bronchopleural complications by active intervention on wound healing (EBLT, spray, drainage of pleural cavity)

		EBLT	Without EBLT	P
N		48	35	
3rd day	Alveolar macrophage	62.35±2.16	60.18±2.17	p<.05
	Neutrophil	19.17±1.5	22.35±1.35	p<.05
	Epithelial cell	13.09±1.04	15.14±1.72	p<.01
	Lymphocyte	4.21±0.32	2.33±0.35	p<.001
	Fibroblast	1.18±0.09	-	-
N		45	30	
8th day	Alveolar macrophage	68.03±2.61	62.92 3.18	p<.05
	Neutrophil	12.25±1.17	20.37±1.94	p<.001
	Epithelial cell	10.22±1.21	14.24±1.11	p<.01
	Lymphocyte	5.38±0.39	2.47±0.16	p<.001
	Fibroblast	4.12±0.57	-	-
N		40	30	
15th day	Alveolar macrophage	70.05±2.74	68.03±3.07	p<.05
	Neutrophil	7.24±0.94	17.17±1.56	p<.001
	Epithelial cell	9.49±1.28	11.27±1.73	p<.05
	Lymphocyte	8.07±1.01	3.12±0.35	p<.001
	Fibroblast	55.14±0.72	0.31±0.01	p<.001
N		35	20	
25th day	Alveolar macrophage	75.17±3.78	72.15±3.21	p<.05
	Neutrophil	4.14±0.42	11.36±1.07	p<.001
	Epithelial cell	7.21±1.18	8.40±0.98	p<.05
	Lymphocyte	6.37±0.78	5.05±0.47	p<.001
	Fibroblast	77.11±1.05	3.05±0.12	p<.001

After all above mentioned it is possible to make early prognosis of complications by observing recovering process using BS cytological investigation. It gives opportunity to decrease frequency of postoperation bronchopleural complications among 1st group patients (50) revealed in 6 (12%) cases while among 2nd group patients (65) - in 3 (4.6%) cases; these data are sufficiently low compared to common statistical data (8-18%).

Finally we can conclude that cytological investigation of specimens obtained by BAL from stump area is a valuable procedure for early diagnostic of bronchopleural complications after pulmonectomy and for observation purposes on recovering process in BS. We also recommend FFB with TBD and EBLT along with routine treatment according to protocol given in article for acceleration of recovering process and diminishing postoperation complications. It will improve effectiveness of pulmonectomy, increase activity of surgical intervention and decrease the number of epidemiologically most dangerous group of patients.

G.Mestiashvili's Memorial
Phthisio-Pulmonologic Hospital

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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N.Chichinadze, M.Bzishvili, M.Kakabadze

Autoradiographic Indicator of Androgen Distribution in the Heart Muscle

Presented by Academician N.Javakhishvili, November 28, 1996

ABSTRACT. **With the use of autoradiographic method, it was observed less intensive concentration of ^{3}H Methyltrienolon in all parts of cardiomyocyte nuclei of intact male and female rats. In myocyte nuclei of heart muscle of castrated male rats the intensive turn on of ^{3}H-MTR was noticable.**

In control castrated animals, first injected with unmarked, then with marked ^{3}H-MTR less concentration of radioactive particles in cardiomyocyte nuclei was noted. Fusion of pituitrin and adrenaline with castration was provoking not only intensive accumulation of radioactivity, but also formulation of focal necroses in myocardium.

The purpose of our investigation was identification of androgen target cells and establishment of topography in various regions of the heart muscle (heart auricles, right and left ventricles and atrium).

Experiment was carried on the following groups of 80 white rats of both sexes (by mass 180-200 g). Group 1 - intact animals, intravenously injected with methyltrienolon (^{3}H-MTR). Group 2 - male castrated rats on the 4th day of castration intravenously injected with ^{3}H-MTR. Group 3 - male castrated rats, on the 4th day of castration intravenously injected first time with unmarked, then in 5 minutes with marked ^{3}H-MTR. Group 4 - male castrated rats, on the 4th day of castration intravenously injected first with both adrenaline and pituitrin, (adrenaline - 20 g on each animal, pituitrin - 0.5 unit), then in 30 minutes with ^{3}H-MTR. Group 5 - male intact rats, intravenously injected with both adrenaline and pituitrin, and in 30 minutes with ^{3}H-MTR. Group 6 - female intact rats, intravenously injected with ^{3}H-MTR. Group 7 - female intact rats, intravenously injected first with unmarked, then in 5 minutes with marked ^{3}H-MTR. Group 8 - female intact rats, intravenously injected with pituitrin and adrenaline and in 30 minutes with ^{3}H-MTR. In each group under test there were 10 rats.

One hour later after androgen injection, the animals were killed by decapitation. Pieces of the heart muscle were formulated in epone. Half thin sheets were coloured with toluidine blue and were covered with emulsion of thin particles. M-exposition lasted for 6 months. Ultrathin sheets were studied by the electron microscope Numerous material was elaborated by the use of statistic methods.

Being injected with ^{3}H-MTR in all parts of cardiomyocyte nuclei of intact male rats, less intensive concentration of radioactive particles, (heart auricle M=25.8; atrium M=24.3 and ventricles M=27.7) was denoted.

In cardiomyocyte nuclei of castrated rats, a large concentration of marked particles was seen in the heart auricle M=57.2, in atrium and ventricles. This radioactivity was properly equal to M=54.1 and M=49.3.

In control castrated animals, first injected with unmarked, then with marked ^{3}H-MTR, concentration of radioactive particles was the following: in the heart auricle

M=7.5, atrium M=8.3 ventricles M=10.6. We observe unmarked MTR blocks inclusion of marked androgens in the nuclei.

Intensive concentration of marked particles with radioactive substances was noted in these castrated rats, whose heart muscle was under pituitrin and adrenaline influence.

The concentration of radioactive particles in myocyte nuclei of ventricles was much more (M=66.3) than in atrium (M=63.3) and in heart auricle (M=61.3).

In some regions of myocardium, the structure of mitochondria was almost disappearing. Protective envelope broke and destruction of myofibrilation and lysis was noted.

Pituitrin and adrenaline activizes a number of androgens in cardiomyocyte nuclei, which is proved by the increasing of the concentration of ^{3}H-MTR in the heart muscle of intact rats (heart auricle M=46.2, atrium M=43.2 and ventricles M=48.1). Concentration of radioactive particles in cardiomyocyte nuclei of female intact rats heart auricle M=60.2; atrium M=61.2 and ventricles M=64.2 increased. The same reaction was observed in male rats after injection of unmarked MTR blocks inclusion of marked MTR in nuclei was the following: heart auricle M=17.9; atrium M=18.5 and ventricles M=18.9.

High loading with pituitrin and adrenaline in rats increases androgen concentration especially in cardiomyocytes of female rats (heart auricle M=68.1; atrium M=86.6 and ventricle M=76.3).

Intensive inclusion of ^{3}H-MTR in cardiomyocyte nuclei of castrated rats can be explained: 1) by decreasing of androgen concentration in blood; 2) by releasing of adrenal glands from androgen inhibition action and hormone gonadotrophic stimulation after castration.

On th 4th day of castration, dehidroandrosterone, produced by the adrenal glands can not supply the heart muscle with sufficient quantity of androgens. Newly injected ^{3}H-MTR can easly influence on myocyte nuclei.

The experiments showed androgen concentration increasing and development of changes in cardiomyocytes of castrated rats due to loading with Pituitrin and Adrenaline. These changes can be explained by energosupplying disbalance between demands and possibilites of the heart muscle hyperfunction which causes ischemia.

The fact of androgen inclusion in female rats heart muscle and especially in ventricles due to action of pituitrin and adrenaline can be explained by the existence of receptor system for androgens in the ventricles of female rats. The heart muscle in the extremal situation needs higher metabolic activity. The direct and indirect action of androgens in ventricles, specific induction of receptors take place in cardiomyocytes.

Tbilisi State Medical University

Georgian Academy of Sciences
A.Natishvili Institute of Experimental
Morphology

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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LINGUISTICS

N.Barnovi

The Structure of Consonant Clusters in the Didoian Languages

Presented by Academician T.Gamkrelidze, November 11, 1996

ABSTRACT. **The paper deals with the structure of consonant clusters in languages of the Didoian subgroup.**

Consonant clusters occur mainly in the inlaut, the majority being two membered. The consonant clusters of the Cezian languages can be subdivided into two classes:

1. With a resonant consonant as its first member; 2. With noisy consonant in the initial position. The second member of a two-consonant cluster in both classes in root morphemes is always a nonlateral, nonpharyngeal, nonlaryngeal noisy consonant.

The structure of consonant clusters has been thoroughly studied in the linguistic literature. However, structural relations (viz. paradigmatic) of the Avarian-Andian-Didoian group's consonant phonemes have not been specifically described.

As it is known, the languages of the Didoian subgroup are: Didoian (Cezian), Hinukhian (although some authors consider it to be a dialect of the Didoian), Khvarshian and Hunzibian (Bezhian).

The consonantism of the Didoian languages is characterized by a greater simplicity than the Avarian-Andian languages, mainly because the Cezian languages do not have the correlation of "strong-week". According to T.Gudava [1] this is caused by the elimination of the opposition of the long and short consonants in the parent language in the period of its division into separate dialects. The dezabruptivation of the long affricates and voicing of the voiceless spirants is one of the main historical regularities of the consonantism of the Didoian languages. As mentioned, the consonant system of modern Cezian languages is shared by all languages and dialects of the group, therefore it is regarded as a "complete system".

In the opinion of specialists, as a result of the gradual loss of the correlation on the "intensivity" in the Cezian languages, instead of the complicated consonant system a fairly simple system has emerged, which stands very near to the consonant phonologic models of the Kartvelian and Nakhian languages.

The main part of our research is dedicated to the investigation of consonant disrtibution in the root.morpheme structures. At the same time we are reviewing the consonant distribution in the clusters that appear only at the juncture of the morphemes and in the clusters that appear as a result of the interference. There are two categories of the consonant groups: 1.Marked groups - the consonant combinations that we might not encounter due to the lack of the available material; 2.Unmarked groups - the consonant combinations, the absence of which is conditioned by the phonological structure of the language (according to G.Machavariani's definition).

In the Didoian languages generally the consonant groups are to be found in the position between the vowels (resp. in inlaut). Therefore our research deals with the consonant groups given in inlauts: In the initial position of the word only the C+W type clusters are present. (This happens very rarely); and in the auslaut position the consonant groups are virtually absent.

The distributional analysis of the data has shown that in the Didoian languages, as well as in other Daghestan languages, the majority of the consonant groups consists of two members, the groups with three members appear very rarely. It should be emphasized, that at the present no dictionaries of the Didoian languages are available, except for a small dictionary of the Hunzibian and Bezian, attached to the researches of the mentioned languages and the Bezian-Russian dictionary by M.Khalilov [2]. Fortunately, the specialists of the Didoian languages have identified numerous lexical parallels between the Didoian and Avarian-Andian languages. The main sources of data for our research were the works of E.Bocariov [3], L.Imnaishvili [4], E.Lomtadze [5], G.Madieva [6] and others. We have also used the comparative dictionary of the Daghestan languages by S.Haidakov [7]. For these purposes we have also used T.Gudava's [1] and B.Gigineishvili's [8] monographs and A.Kibrik's and E.Kodzasov's works [9].

In the consonant clusters of the Didoian languages the following strucrural classes of the consonant pairs can be distinguished: 1. Groups with the noisy consonant as the first member; 2. Groups with the resonant consonant as the first member. As for the second consonant of the 1st and 2nd classes, according to the analysis in the proper lexical items, it always appears with consonant: TT, RT.

In case of a single morpheme (resp.root types) the following general regularity is apparent: the second member of the group is presented with a nonlateral, nonpharyngeal, nonlaryngeal consonant, therefore in the root structure we have: I. T+C (nonlateral, nonpharyngeal, nonlaryngeal); II.R+C (nonlateral, nonpharyngeal, nonlaryngeal).

Therefore, in the consonant clusters of the Didoian languages in case of a single morpheme the posretior consonants do not appear, which agrees with the well-known opinion of Ar.Chikobava, that these consonants are secondary in the consonant systems of these languages. [?]

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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HISTORY

I.Menteshashvili

The British Conquest of India as the Process of Contiguity of Western and Eastern Civilizations

Presented by Corr. Member of the Academy L.Chilashvili, November 11, 1996

ABSTRACT. **This article reviews the conquest of India by the British East India Company. Here the referential points are: the regularity and irreversibility of history and also civilization approach to history. The evolution of consciousness is taken as the driving force of the historical process. Proceeding from the above mentioned this research considers the process of conquest not only the example of deviation from the guideline of historical process, but also as the only and natural way of bringing Western and Eastern civilizations into closer ties.**

The activity of British East Company on the Indian subcontinent expands beyond those wars, diplomatic relations and business affairs carried out by it in the 17th-19th centuries. This process must be taken not only as conquest but also as the encounter of the two worlds - East and West. On the other hand this contact caused unity of world's historical process. In order to explain how such great country as India, with its historical heritage turned into dependency of the remote island, which could subjugate it [6] we must reveal that very internal potential which determined their further evolution. Relations between these two countries were determined not only by the ratio of their strength but also by their world vision, values and behaviour accumulated in the course of the centuries [1]. Thus the success of the East India Company in India must be taken not only as the result of her military superiority, but as an output of that preponderance which consists in higher concentration, communicative ability of social organism or its more complex labour activity [5].

Various archive data, published documents, historical and philosophic researches made the sense of my investigation that while contacts between East and West the superiority was on the side of West. Through the differences between the civilizations their contacts turned into relations between the metropolis and the colony. This phenomenon is not only the implementation of the forcible factors within the history and aggregate of precarious elements, but the socio-economic mechanism, based on international division of labour [3]. This turned into base for numerous cultural or social events which caused unity of the world's historical process.

Though West had at his disposal the resources of East still it does not mean it has ceased evolution of Oriental countries. Moreover, it would be an error to prove the hypotethical possibility of the capitalist evolution of Oriental countries earlier than West [3]. It will be better to shed lustre on those spiritual on material factors which troubled separate capitalist evolution of East. This needs at first to clarify the potential of that human staff which comes on historical arena [4]. The capacity of its self-expression depends on the extent of the evolution of the idea of freedom within the concrete country and the extent of assumption of this idea by society itself [2]. The perfection of this process encourages both individual and communal evolution. But this process did not take place in East through the lack of the personal individuality and predominance of the state. The same reasons caused absence of dynamics in the development of Oriental economics. Through the investigation of that socio-economic and cultural environment where the evolution of English and Indian history took place it was established that British conquest of India was natural and irreversible historical phenomenon caused by the requirements of universal historical process.

Georgian Academy of Sciences
Georgian State Museum

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №2, 1996
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HISTORY

K.Kutateladze

From the History of Struggle between Feudal Families in the Second Half of the XII Century

Presented by Academician M.Lortkipanidze, August 19, 1996

ABSTRACT. **In the second half of the XII century the opposition arose between the Orbels and Kvirikians, the inspirator of which was the Georgian King David V. With the purpose to weaken Orbel's family David V, being in conflict with them, took away Amirspasalar's post and gave it to Abuletisdze family. As to Kvirike, he promised him to restore Tashir-Dzoraget Kingdom. The restoration of Kvirikians power in Kvemo Kartli meant leaving the Orbels without lands. David V's politics was dangerous for the Orbels family. That was why they connected with George III and according to Armenian sources (Vardan, Mkhitar Gosh) killed King David. As to Kvirikians, they relieved the tension by dynasty marriage. Submit Orbel's son Ivane married Kvirike III's daughter Rusudan (mentioned Rusugan in Armenian inscriptions) and got political rights to their former property.**

In the middle centuries the resistance between the families or opposition with the king always took place in Georgia if it concerned their dignity, privileges and infringe on their patrimony. The history of Georgia is rich with such facts as well as that of Europe. So we shall discuss the opposition between the two families – the Orbeliani and the Kvirikian, the inspirator of which was Georgian King David the fifth.

The Kvirikians family is the branch of Armenian Bagratuni. Kvemo Kartli lands conquered by the Armenians were given by Shiraki King Ashot III to his son Gurgen, who gave the beginning to the new dynasty and founded in Kvemo Kartly the Tashir-Dzoraget Armenian Kingdom. The name of the family comes from the second name of Gurgen-Kvirike. The Tashir Dzoraget Kingdom existed much longer than the other Armenian kingdoms. In the second half of the XI century it was greatly struck by the Georgian King Bagrat IV, who seizured Samshvilde the capital of the Tashir Dzoraget Kingdom in 1065. The political centre of the Kvirikians moved to Lore. The Kvirikians left Lore oppressed by the Turks at the beginning of the XII century, but they kept the power being the owners of Matsnaberd castle. One branch of their family settled in Norberd castle. From all the other branches of Armenian Bagratuns this branch had the longest history of existence. After Kvirike II death his sons David and Abas became the successors of the throne. At the time of the last Kvirikian kings the kingdom became weak. The famous Armenian historians Kirakos Gandzaketsi and Vardan consider that the main reason of brothers leaving their kingdom were the Georgians [1]. Not counting the occupation of Samshvilde by Bagrat IV the Georgians always struggled in those places against the Turks. The running of David and Abas is mostly happened because of the Turks and not of the Georgians.

In 1104-1105 the owner of Adarbadagan – Mukhamed rose against Sultan Barkhiaruk. Seldzuk state was enveloped by inner wars. At the same time so-called Khzil Amira, who took the advantage of disorder in Iran, occupied Lore and Dwin and killed its owner Abu Nasri – the brother of Manuche Shedadiani [2]. According to

Kirakos Gandzaketsi he destroyed the Sanian and Akhpat monasteries [1]. Khizil Amira is considered Khizil Arslan, who in the war at Khoi fought on Mukhameds side[11]. According to G.Dzaparidze "He as we see took occasion of the destability in this part of Transcaucasia and by no means with Mukhamed's "blessing" put his hand on Lore, and after that took Dwin [11, ebb]. Finally Mukhamed, the Great Seldzuk's sultan did away with Kizil Arslan because of his relations with Kilich Arslan, the Rumi Seldzuk's sultan. Some years later the raid of Seldzuks took place again. Ioan Sarkavag's history, which is represented in "Kiotuk" of Sanaian, contains this notice [12]. The notice is dated by 1110 year according to which David, the son of Kvirike in August of the same year was in Sanaian, when there happened the surprise attack on Sanaian by "Persia" (the seldzuks) [12, 13]. G. Dzaparidze considers that Sultan Mukhamed's supporters occupied Lore by that time [11]. As we see David Kvirikiani was still the owner of Tashir - Dzoraget. But he had to leave his native land together with his brothers. Mkhitar Airivanets names the date of their leaving by 1111 year [3]. According to their notices Matevosian considers that the brothers' leaving had to happen in 1111-1113 years caused by the occupation of their lands by Seldzuks. Thus there was not the Georgians fault in their running. David and Abas running happened before 1118 before David the Restorer appeared in Lore. David took away Lore from the Seldzuks and not from Armenians. The brothers fortified in Tavushi and Matsnaberdi castles. In the middle of XII century Abas lost Tavushi castle and had to pass to his brother to Matsnabers. In 1145 both of them died. David's son Kvirike III took the Kvirikians' power.

As it is generally known the Orbeli family settled in Kvemo Kartli after the Kvirikians running. They took an active part in the liberation of this district. Just because of their selfless and faithful labour David the Restorer gave the Orbels Samshvilde and Lore. So the Orbels seized the residences of the Armenian Kings. They took the honour of Amirspasalar's governing territory. The historian of this family Stephanoz Orbeliani considers that they came from China (Tchenistan, Dzenistan) and helped the Georgians in their struggle against Pershians and for that they were given the Orbeti castle for settling. According to the name of the castle they were called "Orbeliani" i.e. from Orbeti [4]. Stephanoz Orbeliani's version is more praising of the Orbel's family than gives real historical fact. It is clear that the surname Orbeli is connected with the Orbi castle which is the old name of Samshvilde [16]. Stephanos Orbeliani names it "Orbeti". It is mentioned in the same way in the old Armenian translations of Kartlis Tskhovreba [8]. In Armenian historiography the conception exists that "The Orbelians is one of the branches of Mamikonian family which was settled in Gugark" [14]. They consider them to follow the Kalkedonic religion.

The strength of the Orbels family comes from their royal posts, by which they had great influence on royal court. The Orbels possessed the whole war department and hold the post of Mandaturtukhutsesi. Beginning from David the Restorer the Orbels were known as the people faithful to the King. The first tenseness of relations between the Georgian King and the Orbels took place at the time of David V son-Demetre I, who ruled only 6 months. As it is generally known David was in conflict with his father. Tamar's first historian writes that Demetre "preferred younger son and listening to his entreaty God shortened his days and the father (Demetre) brought us his son (George) similar to him and made him the joint owner of the power" [5]. After Lasha-George's historian Demetre took monastic vows in his old age and died in a year". And that was his son David, ruled 6 months and died" [6]. Iv. Dzavakhishvili shared the opinion after Mari Brosset that the fact of taking monastic vows by Demetre was forced, as he was asking the God David's death, knowing that David would have never given him the power [10]. And really the historian of Lasha-George time writes that after having taken monastic vows Demetre died in a year. but as we have seen before, Tamar's historian says that Demetre chose his son George for joint ruling. Iv. Dzavakhishvili concludes that "after taking monastic vows Demetre turned again to the royal life and set his favorite son George III on the throne for joint rulling [10].

The question of the XII century Georgian throne hereditary is not quite clear. David V evidently fights for the throne and by different measures taken it. The team of David's supporters is not known, but one is clear that it must have been rather strong, because in spite of Demetre's sympathy to his younger brother George, David could take the throne. His main supporter was Tirkash' son of Ivane Abuletisdze, who was killed by Demetre in 1132. Tirkash was the king's prisoner. After his accession to the throne, David set him free and appointed him to the Amirspasalar's post. By that he turned all the Amirspasalars of Orbel's family against himself. From Vardan's words because of that the Orbels (Sumbat and Ivane) connected with George III who had promised them to return the Amirspasalar posts and killed David the King [2]. Important notices about David's supporter feudal lords gives us Mkhitar Goshi in his 'Kroniks of Albania". The historian says that David has been on good terms with Vasak Iskhan-the satrap of Tbilisi and his brothers. David was so kindhearted that sent a man and invited Kvirike, King David Bagratuni's son and promised him to return his heredity, taken away from his ancestors. Having presented him in such a way he sent him back and agreed with him about their future meeting [15, p.46-47]. There is almost nothing known about Kvirike III. But this person especially attracts our attention in his connection with the royal court of Georgia in the second part of the XII century. Then historian countries: "when the Georgian nobles learned his such desire, the great envy rose in them, especially in those, who were named as Orbelians; they gave him poison and killed David the King. Hard and long sadness seized the Georgians and the Armenian countries" [15].

As we see David V looked for the numerical strength of his supporters not only in Georgians. With their help he tried to weaken his rival among Georgian nobles. It was known for David the great desire of the Armenians to restore Armenian State system. This desire had always been alive in their minds. In case of realization of this fact by the Georgian king's concept, the idea to restore Armenian State system would become more real. It is clear that David's interest in this question was not caused by his kindheartedness to the Armenians. His interest was from his own interests. It is no more chance that David choose Kvirike III Bagratun-Kvirikiani - the Lord of Matsnaberdi, he was direct heir of royal governors dynasty of Tashir-Dzoraget or Lore. David wanted to restore the kingdom of Lore for the purpose to weaken the Orbels. As to Orbels they took the place of Kvirikian Kings of Tashir-Dzoraget. David V's politics connected with the restorance of Armenian State system especially compromised the Orbels' interests. For a certain time they had lost the post of Amirspasalar and then there was the danger to loose the lands of Lore, because the restorance of former Tashir-Dzoraget kingdom of Kvirikians meant to leave the Orbels without lands. If all this is was happened the opposition between the Kvirikians and the Orbels would have been inevitable. There was no way for the Orbels but killing David (by Mkhitar Goshi). Varden's notice about David's killing by Orbels is confirmed by Mkhitar Goshi.

To avoid the conflict and to firm own position Orbels became related with Kvirikians. Ivane Orbeli married Kvirike III's daughter Rusugan. We suppose that this marriage would have happened after this silent conflict and David's death about in

1156 year. By this marriage the Orbels killed those hopes which had been conceived by the promises of Georgian king to Matsnaberdi lords. The Orbels' family by becoming related with Kvirikians got the political rights to their lands. Ivane Orbeli's wife Rusugan was famous for her religious beliefs and charity. Together with her sisters she took part in Kobery, Akhpat and Sanain monasteries restorations, built by their precursors; for that she used her husband's power and authority "because her husband's (the Great Spasala's Ivan's) power and fear is spread everywhere" [7]. All the Armenians living on the territory of Georgian kingdom relied on him. Rusugan is mentioned on inscriptions of Koberi, Akhpat, Sanaiani, Betania and Green Church. In Georgian inscriptions Rusugan is mentioned as Rusudan. In this respect the inscriptions on the south wall of St. George's small church are especially important, which are restored by V. Silogava in such a way : "Christ the God, who was crucified for us, take care of our owner Rusudan here and there" [9]. The second inscription is much spoiled: "Khoroniconi IV (416-1196y).

God glorify the sole of Great Eristavt-Eristavi here and there. During their being here I had the honour to become their daughter-in-law, the wife of Eristavt-Eristavi's son, the Amirspasalar and Mandaturtukhutsesi – Ivane. Armenian King Kvirike's daughter Rusudan".[9].

This inscription confirms aforesaid. The Orbels and Kvirikians relations are confirmed by epigraphic monuments. At first sight the date of inscription makes us to think that after the Orbels' uprising Rusugan was saved of George's anger, because she seemed to be alive in 1196. But as it is known, after uprising Betania is announced as a royal monastery. That is why the representatives of the protaned family could not be in Betania after the uprising in 1177. Hence, the date of the inscription is doubtful. The date of Rusudan's death is not known, she is burried in Akhpat in Kvirikian's family cemetery. There is an Armenian inscription on her tombstone "Rusugan".

As it seen from Mkhitar Goshi's aforesaid inscription David's death caused a great sorrow in Georgian and Armenian countries. Here, of course, David's supporter Georgian and Armenian nobles are meant. David's supporter Georgian nobles must have been the opponents of the Orbels' family, because having been frightened by into order the resistance with the Kvirikians and reserved their lands and position in the state.

I.Javakhishvili Tbilisi State University

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CONTENTS

MATHEMATICS

CYBERNETICS

PHYSICS

ANALYTICAL CHEMISTRY

ORGANIC CHEMISTRY

PHYSICAL CHEMISTRY

GEOLOGY

PETROLOGY

METALLURGY

HYDRAULIC ENGINEERING

ELECTROTECHNICS

AUTOMATIC CONTROL AND COMPUTER ENGINEERING

BIOCHEMISTRY

ECOLOGY

EXPERIMENTAL MORPHOLOGY

EXPERIMENTAL MEDICINE

PALAEOBIOLOGY

HISTORY OF SCIENCE

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №1, 1996

MATHEMATICS

S.Saneblidze

Lusternik-Schnirelmann Category of a Weak Formal Space

Presented by Corr. Member of the Academy N.Berikashvili, December 26, 1994

ABSTRACT. **The paper presents the Lusternik-Schnirelmann category computed for a class of spaces by the integral cohomology ring.**

The Lusternik-Schnirelmann category of a topological space X, $cat(X)$, is the least integer n such that X has an open cover $n+1$ sets, each of which is contractable in X. Lusternik and Schnirelmann [1] proved that if a smooth real - valued function $M \to R$ on a manifold M is bounded below, then it has at least $cat(M)+1$ critical points. A simple approximation of $cat(X)$ is the *cup - length*, $c(X)$, of the integral cohomology $H(X)$. It is the least integer k such that $a_1 a_2 \ldots a_{k+1} = 0$ for all $a_i \in \tilde{H}(X)$. In general, $cat(X) \geq c(X)$.

Here we describe a class of spaces for which the equality just holds above. Namely for a space X with $c(X) = n$ consider the following conditions:

(A) The diagonal map $X \to \Pi^{n+1} X = X \times \ldots \times X$ is weak formal [2];

(B) The following short sequence is exact

$$0 \to H^i(X; \pi_{i-1}(F_X)) \xrightarrow{u^*} H^i(X; H_{i-1}(F_X))$$

where F_X denotes $n+1$ - fold join of the loop space ΩX of X and u^* is induced by the Hurewicz homomorphism in coefficients.

(C) For any sequence of homomorphisms preserving $H^*(X; R\pi_*(F_X))$

$$f_h^j : H^i(X; R_q H_i(F_X)) \to \oplus_{k=1}^{j} H^{i+k}(X; R_{q+j-k} H_{i+j-1}(F_X)),\ j \geq 2,$$

of the form $f_h^j = h\cup\_$, some $h \in H^*(X; Hom(RH_*(F_X), RH_*(F_X)))$ (the cup product is defined by the evaluation map in coefficients), $RH_*(F_X)$ is a free group resolution of $H_*(F_X)$, and for elements $a^{(n)} = (a^1, \ldots, a^n)$, $a^i \in H^i(X; R_0 H_i(F_X)) \oplus H^{i+1}(X; R_1 H_i(F_X))$ and $c^{(n-1)} = (c^1, \ldots c^{n-1})$, $c^i \in H^i(X; R_0 \pi_i(F_X)) \oplus H^{i+1}(X; R_1 \pi_i(F_X))$ with $[f_h(a^{(i-1)} - c^{(i-1)})]^{i+1} = 0$ and $[f_h(a^{(n)})]^{n+2} \in H^{n+2}(X; \pi_{n+1}(F_X))$, $f_h = \left\{ f_h^j \right\}$ there is an element

$$c^n \in H^n(X; R_0 \pi_n(F_X)) \oplus H^{n+1}(X; R_1 \pi_n(F_X)) \text{ such that } [f_h(a^{(n)} - c^{(n)})]^{n+2} = 0.$$

We have the following

Theorem 1. *If a simply connected space X having the homotopy type of a CW - complex satisfies conditions (A), (B) and (C), then*

$$cat(X) = c(X).$$

It is well - known that if $cat(X) = 1$, then X is a co-H-space [3].

Thus, we obtain the following

Corollary 2. *Let $H^*(X)$ have the trivial multiplication. If X satisfies (A) and (B), then X is a co-H-space.*

For example, the hypotheses of the theorem are satisfied for spheres, finite projective spaces, etc.

Now let for a map $g: X \to Y$, $c(g)$ denote the cup - length of the ring $Im(g^*)$, $g^*: H(Y) \to H(X)$ [3].

Then we have the following

Theorem 3. *Let $g:X\to Y$ be a weak formal map with $c(g)=n$. If X, Y satisfy conditions (A), (B') and (C'), where (B') and (C') are obtained respectively from (B) and (C) by replacing F_X by F_Y, then*

$$cat(g)=c(g).$$

The proofs of the theorems are based on an obstruction theory to the section problem in a fibration developed in [2].

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A.Razmadze Mathematical Institute
Georgian Academy of Sciences

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
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MATHEMATICS

M.Ashordia

On the Correctness of the Multipoint Boundary Value Problem for the System of Generalized Ordinary Differential Equations

Presented by Academician I.Kiguradze, April 10, 1995

ABSTRACT. **The sufficient conditions are given to guarantee the correctness of the problem**

$$dx(t)=dA(t)\cdot f(t,x(t)),$$
$$x_i(t_i)=\varphi_i(x)\ (i=1,\dots,n),$$

where $t_1,\dots,t_n\in[a,b]$, $A:[a,b]\to R^{n\times n}$ is a matrix-function with bounded variation components, $f:[a,b]\times R^n\to R^n$ is a vector-function belonging to the Caratheodory class corresponding of A, and $\varphi_i(i=1,\dots,n)$ are the continuous functionals from the space of vector-functions of bounded variation into R.

Let $t_1,\dots,t_n\in[a,b]$; $A_0=(a_{ik0})_{i,k=1}^{n}:[a,b]\to R^{n\times n}$ be a matrix-function with bounded variation components; $a_{ik0}(t)\equiv a_{1ik}(t)-a_{2ik}(t)$, where a_{jik} is a function nondecreasing on the every intervals $[a,t_i[$and$]\ t_i,b]$ for $j\in\{1,2\}$ and $i,k\in\{1,\dots,n\}$;

$$A^{(j)}(t)=(a_{jik}(t))_{i,k=1}^{n}\ (j=1,2);\ f_0=(f_{i0})_{i=1}^{n}:[a,b]\times R^n\to R^n$$

be a vector-function belonging to the Caratheodory class corresponding to matrix-function A_0 and $\varphi_i:BV_s([a,b],R^n)\to R$ $(i=1,\dots,n)$ be a continuous functionals in general nonlinearity.

For the system of generalized ordinary differential equations

$$dx(t)=dA_0(t)\cdot f_0(t,x(t)) \tag{1}$$

consider the multipoint boundary value problem

$$x_i(t_i)=\varphi_i(x)\quad(i=1,\dots,n). \tag{2}$$

Consider a sequence of matrix-functions of bounded variation $A_m:[a,b]\to R^{n\times n}$ $(m=1,2\dots)$, a sequence of vector-functions $f_m=(f_{im})_{i=1}^{n}:[a,b]\times R^n\to R^n$ $(m=1,2,\dots)$ belonging to the Caratheodory class corresponding the matrix-function A_m, a sequence of points $t_{1m},\dots,t_{nm}\in[a,b]$ $(m=1,2,\dots)$ and a sequence of continuos functionals $\varphi_{1m},\dots,\varphi_{nm}:BV_s([a,b],R^n)\to R$ $(m=1,2,\dots)$.

In this paper sufficient conditions are given that guarantee both the solvability of the problem

$$dx(t)=dA_m(t)\cdot f_m(t,x(t)), \tag{1$_m$}$$
$$x_i(t_{im})=\varphi_{im}(x)\quad(i=1,\dots,n) \tag{2$_m$}$$

for any sufficiently large m and convergence of its solutions as $m\to+\infty$ to the solution of the problem (1), (2) if this problem is solvable.

The results of analogous character are contained in [1] for the Cauchy-Nicolletti's boundary value problem ($\varphi_i(x)=c_i$,c_i=const) and in [2,3] for the multipoint boundary value problems for the systems of the ordinary differential equations. The theory of generalized ordinary differential equations enables to investigate the ordinary differential and the different equations from the common standpoint [1, 4-8].

Throughout the paper the following notations and definitions will be used.

$R=]-\infty,+\infty[$, $R_+=[0,+\infty[$; $[a,b]$ $(a,b\in R)$ is a closed segment. $R^{n\times m}$ is a space of all real $n\times m$ - matrices $X=(x_{ij})_{i,j=1}^{n,m}$ with the norm $\|X\|=\max\limits_{j=1,\dots,m}\sum\limits_{i=1}^{n}|x_{i,j}|$; $R_+^{n\times m}=\left\{(x_{ij})_{i,j=1}^{n,m}: x_{ij}\geq 0\ (i=1,\dots,n;\ j=1,\dots,m)\right\}$.

If $X=(x_{ij})_{i,j=1}^{n,m}\in R^{n\times m}$, then $|X|=\left(|x_{ij}|\right)_{i,j=1}^{n,m}$, $[X]_+=\left(\dfrac{|x_{ij}|+x_{ij}}{2}\right)_{i,j=1}^{n,m}$. $R=R^{n\times 1}$ is a space of all real column n-vectors $x=(x_i)_{i=1}^{n}$; $R_+^n=R_+^{n\times 1}$. If $X\in R^{n\times n}$, then X^{-1} and $det(X)$ are respectively the matrix inverse to X and the determinant of X; I is the identity $n\times n$-matrix.

$\overset{b}{\underset{a}{V}}(X)$ is the sum of total variations of components $x_{ij}(i=1,\dots,n;\ j=1,\dots,m)$ of the matrix-function $X:[a,b]\to R^{n\times m}$; $V(X)(t)=(v(x_{ij})(t))_{i,j=1}^{n,m}$, where $v(x_{ij})(a)=0$ and $v(x_{ij})(t)=\overset{t}{\underset{a}{V}}(x_{ij})$ for $a<t\leq b$; $X(t-)$ and $X(t+)$ are the left and the right limit of the matrix-function $X:[a,b]\to R^{n\times m}$ at the point $t(X(a-)=X(a),\ X(b+)=X(b))$. $d_1X(t)=X(t)-X(t-)$, $d_2X(t)=X(t+)-X(t)$; $\|X\|_s=\sup\left\{\|X(t)\|: t\in[a,b]\right\}$.

$BV_s([a,b],R^{n\times m})$ is a normed space of all matrix-functions of the bounded variation $X:[a,b]\to R^{n\times m}$ (i.e., such that $\overset{b}{\underset{a}{V}}(X)<+\infty$) with the norm $\|X\|_s$; $BV_s([a,b],R_+^n)=\{x\in BV_s([a,b],R^n): x(t)\in R_+^n\ (t\in[a,b])\}$. If $y\in BV_s([a,b],R^n)$ and $r\in]0,+\infty[$, then $U(y;r)=\{x\in BV_s([a,b],R^n): \|x-y\|_s<r\}$; $D(y;r)$ is a set of all $x\in R^n$ such that $\inf\left\{\|x-y(\tau)\|: \tau\in[a,b]\right\}<r$.

If $D\subset R$ is an interval, then $C(D,R^n)$ is a set of all continuous vector-functions $x:D\to R^n$; $C(D,R_+^n)=\{x\in C(D,R^n): x(t)\in R_+^n$ for $t\in[a,b]\}$.

If $g:[a,b]\to R$ is nondecreasing function, $x:[a,b]\to R$ and $a\leq s<t\leq b$, then $\int_s^t x(\tau)dg(\tau)=\int_{]s,t[} x(\tau)dg(\tau)+x(t)d_1g(t)+x(s)d_2g(s)$, where $\int_{]s,t[} x(\tau)dg(\tau)$ is the Lebesgue-Stieltjes integral over the open interval $]s,t[$ with respect to the measure μ_g corresponding to the function g (if $s=t$, then $\int_s^t x(\tau)dg(\tau)=0$).

A matrix-function is said to be nondecreasing if every its components are the same.

If $G=(g_{ij})_{i,j=1}^{l,n}:[a,b]\to R^{l\times n}$ is a nondecreasing matrix-function and $D\subset R^{n\times m}$, then $L([a,b], D; G)$ is a set of all matrix-functions $X=(x_{jk})_{j,k=1}^{n,m}:[a,b]\to D$ so that

$$\int_a^b |x_{jk}(t)|\,dg_{ij}(t)<+\infty \ (i=1,\dots,l;\ j=1,\dots,n;\ k=1,\dots,m);$$

$$\int_s^t d\,G(\tau)\cdot X(\tau)=\left(\sum_{j=1}^{n}\int_s^t x_{jk}(\tau)dg_{ij}(\tau)\right)_{i,k=1}^{l,m} \quad \text{for } a\le s\le t\le b.$$

If $D_1\subset R^n$ and $D_2\subset R^{n\times m}$, then $K([a,b]\times D_1,D_2;\ G)$ is the Caratheodory class, i.e., the set of all mappings $F=(f_{jk})_{j,k=1}^{n,m}:[a,b]\times D_1\to D_2$ such that for each $i\in\{1,\dots,l\}$, $j\in\{1,\dots,n\}$ and $k\in\{1,\dots,m\}$: (a) the function $f_{jk}(\cdot,x):[a,b]\to D_2$ is $\mu_{g_{ij}}$-measurable for every $x\in D_1$; (b) the function $f_{jk}(t,\cdot):D_1\to D_2$ is continuous for $\mu_{g_{ij}}$-almost every where $t\in[a,b]$ and $\sup\{\|f_{jk}(\cdot,x)\|:x\in D_0\}\in L([a,b], R_+;\ g_{ij})$ for every compactum $D_0\subset D_1$.

If $G_j:[a,b]\to R^{l\times n}$ $(j=1,2)$ are a nondecreasing matrix-functions, $G=G_1-G_2$ and $X:[a,b]\to R^{n\times m}$, then $\int_s^t dG(\tau)\cdot X(\tau)=\int_s^t dG_1(\tau)\cdot X(\tau)-\int_s^t dG_2(\tau)\cdot X(\tau)$ for $a\le s\le t\le b$; $K([a,b]\times D_1,D_2;G)=\bigcap_{j=1}^{2}K([a,b]\times D_1,D_2;G_j)$.

If $B\in BV([a,b],\ R^n)$, then $M([a,b]\times R_+,\ R_+^n;\ B)$ is the set of all vector-functions $\omega\in K([a,b]\times R_+,\ R_+^n;\ B)$ such that $\omega(t,\cdot)$ is nondecreasing and $\omega(t,0)=0$ for $t\in[a,b]$.

Inequalities between both the vectors and matrices are understood as componentwise.

If B_1 and B_2 are the normed spaces, then an operator $\varphi:B_1\to B_2$ is called positive homogeneous if $\varphi(\lambda x)=\lambda\varphi(x)$ for every $\lambda\in R_+$ and $x\in B_1$. An operator $\varphi:BV_s([a,b],\ R^n)\to R^n$ is called nondecreasing if for every $x,\ y\in BV_s([a,b],\ R^n)$ such that $x(t)\le y(t)$ for $t\in[a,b]$ the inequality $\varphi(x)(t)\le\varphi(y)(t)$ is fulfilled for $t\in[a,b]$.

A vector-function $x\in BV_s([a,b],\ R^n)$ is said to be a solution of the system (1) if

$$x(t)=x(s)+\int_s^t d\,A_0(\tau)\cdot f_0(\tau,x(\tau)) \text{ for } a\le s\le t\le b.$$

Under a solution of the sysem of the generalized ordinary differential inequalities $dx(t)-dA_0(t)\cdot f_0(t,x(t))\le 0\ (\ge 0)$ we understand a vector-function $x\in BV_s([a,b],\ R^n)$ such that

$$x(t)-x(s)-\int_s^t d\,A_0(\tau)\cdot f_0(\tau,x(\tau))\le 0\ (\ge 0) \text{ for } a\le s\le t\le b.$$

Let $l:BV_s([a,b],\ R^n)\to R^n$ be a linear continuous operator and let $l_0:BV_s([a,b],\ R^n)\to R_+^n$ be a positive homogeneous continuous operator. We shall say that a matrix-function $P:[a,b]\times R^n\to R^{n\times n}$ satisfies the Opial condition with respect to the triplet $(l,\ l_0;\ A_0)$ if : (a) $P\in K([a,b]\times R^n,\ R^{n\times n};\ A_0)$ and there exists a matrix-function $\Phi\in L([a,b],\ R_+^{n\times n};\ A_0)$ such that $|P(t,x)|\leq\Phi(t)$ on $[a,b]\times R^n$; (b) $det(I+(-1)^j d_j B(t))\neq 0$ for $t\in[a,b]\ (j=1,2)$ and the problem $dx(t)=dB(t)\cdot x(t),\ |l(x)|\leq l_0(x)$ has only trivial solution for every $B\in BV_s([a,b],\ R^{n\times n})$ provided, there exists a sequence $y_k\in BV_s([a,b],\ R^n)$ $(k=1,2,\ldots)$ such that $\lim\limits_{k\to+\infty}\int_a^t dA_0(\tau)\cdot P(\tau,y_k(\tau))=B(t)$ uniformly on [a,b].

Let x^0 be a solution of the problem (1), (2) and let r be a positive number.

The solution x^0 is said to be a strongly isolated in the radius r if there exist $P\in K([a,b]\times R^n,\ R^{n\times n};\ A_0)$, $q\in K([a,b]\times R^n,\ R^n;\ A_0)$, linear continuous operator $l:BV_s([a,b],\ R^n)\to R^n$, positive homogeneous continuous operator $l_0:BV_s([a,b],\ R^n)\to R_+^n$ and continuous operator $\tilde{l}:BV_s([a,b],\ R^n)\to R^n$ such that: (a) $f_0(t,x)=P(t,x)x+q(t,x)$ for $t\in[a,b]$, $\|x-x^0(t)\|<r$ and the equality $h(x)=l(x)+\tilde{l}(x)$ is fulfilled on $U(x^0;r)$; (b) vector-functions $\alpha(t,\rho)=max\ \{\|q(t,x)\|:\|x\|\leq\rho\}$ and $\beta(\rho)=sup\{[\ |\tilde{l}(x)|-l_0(x)]_+:\|x\|_s\leq\rho\}$ satisfy conditions $\lim\limits_{\rho\to+\infty}\frac{1}{\rho}\int_a^b d(A^{(1)}(t)+A^{(2)}(t))\cdot\alpha(t,\rho)=0$, $\lim\limits_{\rho\to+\infty}\frac{\beta(\rho)}{\rho}=0$; (c) the problem $dx(t)=dA_0(t)\cdot[P(t,x(t))x(t)+q(t,x(t))]$, $l(x)+\tilde{l}(x)=0$ has no solution differing from x^0; (d) the matrix-function P satisfies the Opial condition with respect to the triplet $(l,\ l_0,\ A_0)$.

Let $h(x)=(h_i(x))_{i=1}^n$, $h_i(x)=x_i(t_i)-\varphi_i(x)$ $(i=1,\ldots,n)$ and $h_m(x)=(h_{im}(x))_{i=1}^n$, $h_{im}(x)=$ $=x_i(t_{im})-\varphi_{im}(x)$ $(i=1,\ldots,n;\ m=1,2,\ldots)$ for $x=(x_i)_{i=1}^n\in BV_s([a,b],\ R^n)$. By $W_r(A_0,f_0,\ h;\ x_0)$ we denote a set of all sequences $(A_m,\ f_m,\ h_m)$ $(m=1,2,\ldots)$ such that: (a) $\lim\limits_{m\to+\infty}\int_a^t dA_m(\tau)\cdot f_m(\tau,x)=\int_a^t dA_0(\tau)\cdot f_0(\tau,x)$ uniformly on $[a,b]$ for every $x\in D(x^0;r)$; (b) $\lim\limits_{m\to+\infty}h_m(x)=h(x)$ uniformly on $U(x^0;r)$; (c) there exists a sequence $\omega_m\in M([a,b]\times R_+,\ R_+^n;\ A_m)$ $(m=1,2,\ldots)$ such that $sup\left\{\left\|\int_a^b dV(A_m)(t)\cdot\omega_m(t,r)\right\|:m=1,2,\ldots\right\}<+\infty$,

$$\lim_{s\to 0+}sup\left\{\left\|\int_a^b dV(A_m)(t)\cdot\omega_m(t,s)\right\|:m=1,2,\ldots\right\}=0$$ and $\|f_m(t,x)-f_m(t,y)\|\leq\omega_m(t,\|x-y\|)$ on $[a,b]\times D(x^0;r)$ $(m=1,2,\ldots)$.

The problem (1), (2) is said to be $(x^0;r)$-correct if for every $\varepsilon\in]0,r[$ and $((A_m,\ f_m,\ h_m))_{m=1}^{+\infty}\in W_r(A_0,f_0,\ h,\ x^0)$ there exists a natural m_0 such that the problem (1_m), (2_m) has at least one solution containing in $U(x^0;r)$ and every such solution belongs to the ball $U(x^0;\varepsilon)$ for any $m\geq m_0$.

The problem (1), (2) is said to be correct if it has a unique solution x^0 and for every $r>0$ it is $(x^0;r)$-correct.

We shall say that a pair $((c_{il})_{i,l=1}^{n};(\varphi_{0i})_{i=1}^{n})$ consisting of the matrix-function $(c_{il})_{i,l=1}^{n}\in BV_s([a,b],\ R^{n\times n})$ and of the positive homogeneous nondecreasing operator $(\varphi_{0i})_{i=1}^{n}:BV_s([a,b],\ R_+^n)\to R_+^n$ belongs to the set $U(t_1,\dots,t_n)$ if functions $c_{il}(i\neq l;\ i,\ l=1,\dots,n)$ are nondecreasing on $[a,b]$ and continuous at the point t_i, $d_jc_{ii}(t)\geq 0$ for $t\in[a,b]$ $(j=1,2;\ i=1,\dots,n)$ and the problem

$$[dx_i(t)\text{-}sign(t-t_i)\sum_{l=1}^{n}x_l(t)dc_{il}(t)]sign(t-t_i)\leq 0\ (i=1,\dots,n),$$
$$(-1)^jd_jx_i(t_i)\leq x_i(t_i)d_jc_{ii}(t_i)\ (j=1,2;\ i=1,\dots,n);$$
$$x_i(t_i)\leq\varphi_{0i}(|x_1|,\dots,|x_n|)\ (i=1,\dots,n)$$

has no nontrivial, non-negative solution.

Theorem 1. *Let conditions* $(-1)^{\sigma+1}f_{k0}(t,x_1,\dots,x_n)sign[(t-t_i)x_i]\leq\sum_{l=1}^{n}p_{\sigma ikl}(t)|x_l|+q_k(t,\|x\|)$

for $\mu_{a_{\sigma ik}}$ *- almost everywhere* $t\in[a,b]\setminus\{t_i\}$ $(i,k=1,\dots,n)$,

$$[(-1)^{\sigma+j+1}f_{k0}(t_i,\ x_1,\dots,x_n)signx_i\text{-}\sum_{l=1}^{n}\alpha_{\sigma ikjl}|x_l|\text{-}q_k(t_i,\|x\|)]d_ja_{\sigma ik}(t_i)\leq 0\ (j=1,2;\ i,k=1,\dots,n)$$

be fulfilled on R^n *for every* $\sigma\in\{1,2\}$ *and let inequalities*

$$|\varphi_i(x_1,\dots,x_n)|\leq\varphi_{0i}(|x_1|,\dots,|x_n|)+\gamma\left(\sum_{l=1}^{n}\|x_l\|_s\right)\ (i=1,\dots,n)$$

be fulfilled on $BV_s([a,b],\ R^n)$, *where* $\alpha_{\sigma ikjl}\in R$ $(j,\sigma=1,2;\ i,\ k,\ l=1,\dots,n)$, $(p_{\sigma ikl})_{k,l=1}^{n}\in L([a,b],\ R^{n\times n};\ A^{(\sigma)})$ $(\sigma=1,2;\ i=1,\dots,n)$, $q=(q_k)_{k=1}^{n}\in K([a,b]\times R_+,\ R_+^n;\ A^{(\sigma)})$ $(\sigma=1,2)$ *is a vector-function nondecreasing with respect to the second variable,* $\gamma\in C(R_+,R_+)$ *and*

$$\lim_{\rho\to+\infty}\frac{1}{\rho}\int_a^b d(A^{(1)}(t)+A^{(2)}(t))\cdot q(t,\rho)=0,\quad \lim_{\rho\to+\infty}\frac{\gamma(\rho)}{\rho}=0.$$

Moreover, let there exist a matrix-function $(c_{il})_{i,l=1}^{n}\in BV_s([a,b],\ R^{n\times n})$ *such that*

$$(c_{il})_{i,l=1}^{n};\ (\varphi_{0i})_{i=1}^{n}\in U(t_1,\dots,t_n),\tag{3}$$

$$\sum_{\sigma=1}^{2}\sum_{k=1}^{n}\int_s^t p_{\sigma ikl}(\tau)da_{\sigma ik}(\tau)\leq c_{il}(t)-c_{il}(s)$$

for $a\leq s\leq t<t_i$ *and* $t_i<s<t\leq b$ $(i,l=1,\dots,n)$, (4)

$$\sum_{\sigma=1}^{2}\sum_{k=1}^{n}\alpha_{\sigma ikjl}\,d_ja_{\sigma ik}(t_i)\leq d_jc_{il}(t_i)\ (j=1,2;\ i,l=1,\dots,n).\tag{5}$$

If the problem (1), (2) *has no more one solution, then it is correct.*

Theorem 2. *Let conditions* $(-1)^{\sigma+1}[f_{k0}(t,x_1,\dots,x_n)\text{-}f_{k0}(t,y_1,\dots,y_n)]sign[(t-t_i)(x_i\text{-}y_i)]\leq\sum_{l=1}^{n}p_{\sigma ikl}(t)|x_l\text{-}y_l|$

for $\mu_{a_{\sigma ik}}$*-almost everywhere* $t\in[a,b]\setminus\{t_i\}$ $(i,k=1,\dots,n)$,

$$\{(-1)^{\sigma+j+1}[f_{k0}(t_i,x_1,\dots,x_n)-f_{k0}(t_i,y_1,\dots,y_n)]sign(x_i-y_i)-\sum_{l=1}^{n}\alpha_{\sigma ik\,jl}|x_l-y_l|\}d_j a_{\sigma ik}(t_i)\le 0$$

$$(j=1,2;\ i,k=1,\dots,n)$$

be fulfilled on R^n *for every* $\sigma\in\{1,2\}$, *and let the inequalities* $|\varphi_i(x_1,\dots,x_n)-\varphi_i(y_1,\dots,y_n)|\le\varphi_{0i}(|x_1-y_1|,\dots,|x_n-y_n|)$ $(i=1,\dots,n)$ *be fulfilled on* $BV_s([a,b],R^n)$, *where* $\alpha_{\sigma ikjl}\in R$ $(j,\sigma=1,2;\ i,k,l=1,\dots,n)$, $(p_{\sigma ikl})_{k,l=1}^{n}\in L([a,b],R^{n\times n};A^{(\sigma)})$ $(\sigma=1,2;\ i=1,\dots,n)$. *Moreover, let there exist a matrix-function* $(c_{il})_{i,l=1}^{n}\in BV_s([a,b],R^{n\times n})$ *such that conditions* (3)-(5) *hold. Then the problem* (1), (2) *is correct.*

Ministry of Defence of Georgia

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MATHEMATICS

M.Bakuridze

Some Properties of Fourier Trigonometric Series of Even and Odd Functions

Presented by Academician L.Zhizhiashvili, June 20, 1995

ABSTRACT. **Problems of continuity of sums of definite type "cosine" and "sine" series are considered. The criterion of the sum of "sine" series from the class H^{ω} is established.**

1. Let $f\in L(T)$ be an even function and $g\in L(T)$ be an odd function. Let Fourier series of these functions be, correspondingly

$$\sigma[f]=\sum_{K=1}^{\infty} a_K \cos Kx \quad \text{and} \quad \sigma[g]=\sum_{K=1}^{\infty} b_K \sin Kx.$$

Together with these series we consider the following series [2-4]

$$\sum_{K=1}^{\infty} U_K \cos Kx, \tag{1}$$

$$\sum_{K=1}^{\infty} U_K \sin Kx, \tag{2}$$

$$\sum_{K=1}^{\infty} U_K^* \cos Kx, \tag{3}$$

$$\sum_{K=1}^{\infty} U_K^* \sin Kx, \tag{4}$$

where $U_K=\dfrac{1}{K}\sum_{j=1}^{K} a_j$ and $U_K^*=\sum_{j=K}^{\infty}\dfrac{a_j}{j}$.

Also, we consider the series

$$\sum_{K=1}^{\infty} M_K \cos Kx, \tag{5}$$

$$\sum_{K=1}^{\infty} M_K \sin Kx, \tag{6}$$

$$\sum_{K=1}^{\infty} M_K^* \cos Kx, \tag{7}$$

$$\sum_{K=1}^{\infty} M_K^* \sin Kx, \tag{8}$$

where

$$M_K = \frac{1}{K}\sum_{j=1}^{K} b_j \quad \text{and} \quad M_K^{*} = \sum_{j=K}^{\infty} \frac{b_j}{j}.$$

Let ω be the modulus of continuity [4]. It is said, that ω satisfies Bari-Stechkin's condition [1], if

$$\int_0^{\sigma} \frac{\omega(t)}{t} dt + \sigma \int_{\sigma}^{\pi} \frac{\omega(t)}{t^2} dt = O(\omega(\sigma)), \qquad \sigma \to t$$

In the sequel, if $f \in C(t)$ then by $\omega(\sigma, t)_C$ we denote the modulus of continuity of the function f, and by H^{ω} - the following class of functions:

$$H^{\omega} = \{f: f \in C(t), \quad \omega(\sigma, t)_C = O(\omega(\sigma))\}$$

The theorems stated in the paper concern the behaviour of (1) - (8) series.

Theorem 1. *Let $f \in C(T)$ be an even function and $a_k \geq 0$ $(k = 1, 2, \ldots)$. Then the sums of the series* (3) *and* (4) *are continuous and the sums of the series* (1) *and* (2) *are not generally continuous.*

Theorem 2. *Let $g \in C(T)$ be an odd function and $b_k \geq 0$. Then the sum of the series* (8) *is continuous function and the sums of the series* (5), (6) *and* (7), *are not generally continuous.*

Theorem 3. *Let the modulus of continuity ω satisfies Bari - Stechkin condition* (9), *$g \in C(T)$ be an odd function and $b_k \geq 0$ $(k = 1, 2, \ldots)$. Then, the necessary and sufficient condition for the sum of the series* (8) *to be of H^{ω} class is*

$$k \cdot M_k^{*} \leq A \cdot \omega\left(\frac{1}{k}\right), \; k \geq 1.$$

I.Javakhishvili Tbilisi State University

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MATHEMATICS

T.Akhobadze

On Some Analogies of Riemann-Lebesgue Theorem for Functions of Multiple Variables - I

Presented by Academician L.Zhizhiashvili, June 21, 1995

ABSTRACT. **In the present paper the limits in Pringsheim's sense of the type**

$$\lim_{\lambda_1,\lambda_2\to+\infty}\iint_I f(u,v)\alpha(\lambda_1 u,\ \lambda_2 v)dudv$$

for functions f and α from various classes are investigated.

The Riemann-Lebesgue theorem asserts that Fourier coefficients C_ν of integrable functions concerning the trigonometric system tend to zero as $|\nu|\to\infty$. There exist some generalizations of this theorem. Kahane [2] continued investigations in this direction. In particular he considered the problem of evaluating limits of the type

$$\lim_{\lambda\to+\infty}\int_I f(t)\beta(\lambda t)dt \qquad (1)$$

under various assumptions regarding the functions f and β over the interval I. On the other hand there exist analogies of the Riemann-Lebesgue theorem for multiple cases (i.e. for functions of multiple variables) and some its generalizations.

The purpose of the present paper is to investigate the behaviour of analogy of the limit (1) for multiple case, where the convergence is definite in Pringsheim's sense. From the obtained results in particular, we can get Kahane's [2] corresponding statements and some already known theorems. For the presentation simplicity of our results we shall formulate them for two-dimensional case, although analogies of these statements are valid for arbitrary finite-dimensional cases.

As usual we denote by $L_p\{I\}$ ($1\le p\le+\infty$) the class of Lebesgue integrable to the p-th power functions on the two-dimensional interval, moreover if $p=+\infty$ we designate by $L_\infty\{I\}$ the class of essentially bounded (in Lebesgue's sense) functions on I.

Theorem 1. *Let $\alpha\in L_\infty\{[0,+\infty)^2\}$. Then the necessary and sufficient condition for*

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_0^{+\infty}\int_0^{+\infty} f(u,v)\alpha(\lambda_1 u,\ \lambda_2 v)dudv$$

to exist in Pringsheim's sense for every function $f\in L_1\{[0,+\infty)^2\}$ is existence of the limit

$$M(\alpha)\equiv\lim_{T_1,T_2\to+\infty}\int_0^{T_1}\int_0^{T_2}\alpha(u,v)dudv \qquad (2)$$

also in Pringsheim's sense; moreover this being the case

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_0^{+\infty}\int_0^{+\infty} f(u,v)\alpha(\lambda_1 u,\ \lambda_2 v)dudv=\left(\int_0^{+\infty}\int_0^{+\infty} f(u,v)dudv\right)M(\alpha).$$

Let $\alpha_{11}(u,v)$, $\alpha_{12}(u,v)$, $\alpha_{21}(u,v)$ and $\alpha_{22}(u,v)$ coincide respectively with $\alpha(u,v)$, $\alpha(u,-v)$, $\alpha(-u,v)$ and $\alpha(-u,-v)$ for $u\geq0$ and $v\geq0$.

Corollary 1. *Let* $\alpha\in L_\infty\{(-\infty,+\infty)^2\}$ *then the necessary and sufficient condition for*

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_0^{+\infty}\int_0^{+\infty} f(u,v)\alpha(\lambda_1 u,\ \lambda_2 v)dudv$$

to exist in Pringsheim's sense for every function $f\in L_1\{(-\infty,+\infty)^2\}$ *is that for each function* $\alpha_{ij}\in L_\infty\{[0,+\infty)^2\}$, $i,j=1,2$; *to exist finite limit*

$$\mathrm{M}(\alpha_{ij})=\lim_{T_1,T_2\to+\infty}\frac{1}{T_1T_2}\int_0^{T_1}\int_0^{T_2}\alpha_{ij}(u,v)dudv$$

in Pringsheim's sense. When these mean values exist the following representation is valid.

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_{-\infty}^{+\infty}\int_{-\infty}^{+\infty} f(u,v)\alpha(\lambda_1 u,\lambda_2 v)dudv=\left(\int_0^{+\infty}\int_0^{+\infty} f(u,v)dudv\right)\mathrm{M}(\alpha_{11})+$$

$$+\left(\int_0^{+\infty}\int_{-\infty}^{0} f(u,v)dudv\right)\mathrm{M}(\alpha_{12})+\left(\int_{-\infty}^{0}\int_0^{+\infty} f(u,v)dudv\right)\mathrm{M}(\alpha_{21})+$$

$$+\left(\int_{-\infty}^{0}\int_{-\infty}^{0} f(u,v)dudv\right)\mathrm{M}(\alpha_{22}).$$

It is clear that Theorem 1 and Corollary 1 contain analogy of Riemann-Lebesgue statement for two-dimensional case. In particular from this theorem concerning functions of two variables we can get various analogies for already known statements of Fejer [3], Zygmund [1] and other authors which can be found for example in Polya-Szegö [4].

Let γ be a positive, locally-integrable function on $[0,+\infty)^2$. We say that a function γ belongs to the class $M(\gamma\in M)$ if for every positive constant C there exists a positive constant $M(C)$ such that for all positive T

$$\frac{1}{T}\int_0^{C}\int_0^{T}\gamma(u,v)dudv\leq M(C),$$

$$\frac{1}{T}\int_0^{T}\int_0^{C}\gamma(u,v)dudv\leq M(C). \qquad (3)$$

Theorem 2. *Let* α *be a locally-integrable to the p-th* ($1<p<+\infty$) *power function on the interval* $[0,+\infty)^2$, *moreover suppose that* $|\alpha|\in M$. *Then, in order that the limit*

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int_a^{b}\int_c^{d} f(u,v)\alpha(\lambda_1 u,\lambda_2 v)dudv\quad (0\leq a<b<+\infty,\ 0\leq c<d<+\infty) \qquad (4)$$

to exist in Pringsheim's sense for every function $f\in L_q\{[a,b]\times[c,d]\}$, *where* $\frac{1}{p}+\frac{1}{q}=1$, *it is necessary and sufficient that*

1) *the averages*

$$\frac{1}{T_1T_2}\int\limits_{\frac{a}{b}T_1}^{T_1}\int\limits_{\frac{c}{d}T_2}^{T_2}|\alpha(u,v)|^p\,dudv$$

be bounded as $T_1,T_2\to+\infty$, and

2) *to exist limit* (2) *in Pringsheim's sense. The conditions* 1) *and* 2) *holding then*

$$\lim_{\lambda_1,\lambda_2\to+\infty}\int\limits_a^b\int\limits_c^d f(u,v)\alpha(\lambda_1 u,\lambda_2 v)dudv=\left(\int\limits_a^b\int\limits_c^d f(u,v)dudv\right)M(\alpha). \quad (5)$$

Remark 1. If in the Theorem 2 a=0 and c=0, then the condition $|\alpha|\in M$ is unnecessary. If a or c are positive then the condition $|\alpha|\in M$ in the definite sense can't be weakened.

Theorem 3. *Let α be locally-integrable function on* $[0,+\infty)^2$. *Suppose that the function $\gamma(u,v)=|\alpha(u,v)|$ satisfies the condition* (3). *Then the necessary and sufficient condition for finite limit* (4) *to exist in Pringsheim's sense for every essentially bounded function f is that*

1) $$\frac{1}{T_1T_2}\int\limits_0^{T_1}\int\limits_0^{T_2}|\alpha(u,v)|dudv$$

be bounded as $T_1,T_2\to+\infty$, and 2) *to exist limit* (2) *in Pringsheim's sense. If the conditions* 1) *and* 2) *are fulfilled then* (5) *is valid.*

Remark 2. If in the Theorem 3 a=0 and c=0 then the condition (3) for $\gamma(u,v)=|\alpha(u,v)|$ function are unnecessary too. If a or c are positive then the condition $|\alpha|\in M$ can't be weakened.

Remark 3. The Theorem 3 is new in one-dimensional case too, moreover in this case condition (3) are unnecessary.

I.Javakhishvili Tbilisi State University

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MATHEMATICS

I.Khasaia

Problems of Convergence Multiple Expansions in Complex Sphere for Polynomial Differential Bundle with Constant Coefficients

Presented by Academician L.Zhizhiashvili, December 20, 1995

ABSTRACT. **In case of constant coefficiental operator of the third order the problem of convergence spectral expansion has been studied in complex domain. The analogue of Abeli's well-known theorem about convergence power series has been obtained.**

Let's discuss differential bundle made by the differential equation

$$y^{(3)} - a\lambda y^{(2)} + a\lambda^2 y^{(1)} - \lambda^3 y = 0 \qquad (1)$$

and with expanded boundary conditions

$$y(0) = y'(0) = y(1) = 0 \quad (a = 1 - \sqrt{2}). \qquad (2)$$

Roots of characteristic equation are indicated with $\alpha_1, \alpha_2, \alpha_3$.

Suppose they are numbered so that $\alpha_1 = 1$, $\alpha_2 = e^{i\theta}$, $\alpha_3 = e^{-i\theta}$, where $\theta = 3/4\pi$.

The fundamental system of solution of equation (1) has the following form

$$y_j = e^{\lambda \alpha_j x}, \ 1 \le j \le 3 \ . \qquad (3)$$

λ-plane is divided into such S_j $(0 \le j \le 5)$ sectors that

$$\mathrm{Re}\lambda\alpha_1 \ge \mathrm{Re}\lambda\alpha_2 \ge \mathrm{Re}\lambda\alpha_3, \ \lambda \in S_0$$
$$\mathrm{Re}\lambda\alpha_1 \ge \mathrm{Re}\lambda\alpha_3 \ge \mathrm{Re}\lambda\alpha_2, \ \lambda \in S_1$$
$$\mathrm{Re}\lambda\alpha_3 \ge \mathrm{Re}\lambda\alpha_1 \ge \mathrm{Re}\lambda\alpha_2, \ \lambda \in S_2$$
$$\mathrm{Re}\lambda\alpha_3 \ge \mathrm{Re}\lambda\alpha_2 \ge \mathrm{Re}\lambda\alpha_1, \ \lambda \in S_3$$
$$\mathrm{Re}\lambda\alpha_2 \ge \mathrm{Re}\lambda\alpha_3 \ge \mathrm{Re}\lambda\alpha_1, \ \lambda \in S_4$$
$$\mathrm{Re}\lambda\alpha_2 \ge \mathrm{Re}\lambda\alpha_1 \ge \mathrm{Re}\lambda\alpha_3, \ \lambda \in S_5.$$

The following lemma takes place.

Lemma 1. For any $\lambda \in S_j$ $(0 \le j \le 5)$ the following unequalities take place

$$\frac{\theta}{2}(j - 1) \le \arg\lambda \le \frac{\theta}{2}j, \ j = 0, 1$$

$$\frac{\pi}{2}(j - 2 - 2\delta_{j,5}) + \frac{\theta}{2} \le \arg\lambda \le \pi(j - 1 - 2\delta_{j,5}) - \frac{\theta}{2}, \ j = 2, 5$$

$$\pi - \frac{\theta}{2}(j - 2 - 2\delta_{j,4}) \le \arg\lambda \le \pi + \frac{\theta}{2}(j - 3), \ j = 3, 4 ,$$

where $\delta_{j,k}$-is the symbol of Cronecker.

It isn't difficult to notice that the equation has the form for its eigen numbers [1]

$$\Delta(\lambda) = 0 ,$$

where

$$\Delta(\lambda) = \lambda \sum_{j=1}^{3} e^{\lambda \alpha_j} \Omega_j$$

and

$$\Omega_1 = \begin{vmatrix} 1 & 1 \\ \alpha_2 & \alpha_3 \end{vmatrix}, \quad \Omega_2 = \begin{vmatrix} 1 & 1 \\ \alpha_3 & \alpha_1 \end{vmatrix}, \quad \Omega_3 = \begin{vmatrix} 1 & 1 \\ \alpha_1 & \alpha_2 \end{vmatrix}$$

Theorem 1. (1)-(2) boundary problem has three infinite sequence of eigen numbers, that satisfy the following asymptotic

$$\lambda_{k,1} = \frac{\pi + 2\pi k - i\ln b_1}{2\sin\frac{\theta}{2}} e^{-\frac{\theta}{2}i} + O(\frac{1}{k}) = \lambda^{\circ}_{k,1} + O(\frac{1}{k}),$$

$$\lambda_{k,2} = \frac{-\pi + 2\pi k - i\ln b_2}{2\sin\frac{\theta}{2}} e^{\frac{\theta}{2}i} + O(\frac{1}{k}) = \lambda^{\circ}_{k,2} + O(\frac{1}{k}),$$

$$\lambda_{k,3} = \frac{-\pi + 2\pi k - i\ln b_1}{2\sin\theta} e^{\pi i} + O(\frac{1}{k}) = \lambda^{\circ}_{k,3} + O(\frac{1}{k}),$$

where b_j $(1 \le j \le 3)$ are to λ independent numbers.

Let's take function

$$\psi(x, \lambda) = \lambda^{-1} \varphi(x, \lambda),$$

where

$$\varphi(x, \lambda) = \begin{vmatrix} y_1(0) & y_2(0) & y_3(0) \\ y'_1(0) & y'_2(0) & y'_3(0) \\ y_1(x, \lambda) & y_2(x, \lambda) & y_3(x, \lambda) \end{vmatrix}$$

$\psi(x, \lambda)$ is eigen function that is corresponding $\lambda = \lambda_k$ to eigen number (λ_k, $k \in N$ eigen numbers are counted according the growth of the model).

From here and (3) we get

$$\psi(x, \lambda) = \sum_{j=1}^{3} e^{\lambda \alpha_j x} \Omega_j .$$

The following lemmas are proved using corresponding questions from [2,3].

Lemma 2. If $\gamma_p^{(i)} \le \arg z \le \delta_p^{(i)}$ $(1 \le i \le 3,\ 1 \le p \le 3)$,

where $\gamma_1^{(i)} = 0,\ \gamma_{p+1}^{(i)} = \delta_p^{(i)} = \dfrac{(5+\delta_{i,3})p - i + 2}{8}\pi,\ 1 \le p \le 2,\ \delta_3^{(i)} = 2\pi,\ 1 \le i \le 3,$

then

$$\left|\psi(z, \lambda_{k,i})\right| \le \begin{cases} C\left|e^{\lambda_{k,i}\alpha_1 z}\right|, & i = p,\ 1 \le p \le 3 \\ C\left|e^{\lambda_{k,i}\alpha_{6-p-i} z}\right|. & 3 - i \le p \le 3,\ 3 - i + 1 \le p \le 2 \end{cases}$$

Constant C does not depend upon k and p.

Lemma 3. *Suppose* $z_0 = |z_0|e^{\theta_0 i}$,

where $\theta_0 \in (\gamma_q^{(i)}, \delta_q^{(i)})$, $1 \le q \le 3$, $\gamma_1^{(i)} = 0$, $\gamma_{q+1}^{(i)} = \delta_q^{(i)} = \dfrac{(5+\delta_{i,3})q - i + 2}{8}\pi$

$\delta_3^{(i)} = 2\pi$, $(1 \le i \le 3)$.

Then

$$\left|\psi(z_0, \lambda_{k,i})\right| \ge \begin{cases} C\left|e^{\lambda_{k,i}\alpha_i z_0}\right|, & i = q, (1 \le p \le 3) \\ C\left|e^{\lambda_{k,i}\alpha_{6-q-i} z_0}\right|, & \text{the rest for } q. \end{cases}$$

Lemma 4. *Suppose that* $[c, d]$ *is an arbitrary segment*

$$c = |c|e^{\theta_0 i},\ d = |d|e^{\theta_0 i},\ 0 < |c| < |d| < 1.$$

Then K_0 *is dependent upon* $[c, d]$ *that* $\forall k > k_0$, $\exists z_k \in [c, d]$.

Then

1) $\lambda_k = \lambda_{k,1}$

$$\left|\psi(z_k, \lambda_{k,1})\right| \ge \begin{cases} C\left|e^{\lambda^0_{k,1}\alpha_1 z_k}\right|, & \theta_0 = 0, \frac{3}{4}\pi \\ C\left|e^{\lambda^0_{k,1}\alpha_3 z_k}\right|, & \theta_0 = \frac{11}{8}\pi \end{cases}$$

2) $\lambda_k = \lambda_{k,2}$

$$\left|\psi(z_k, \lambda_{k,2})\right| \ge \begin{cases} C\left|e^{\lambda^0_{k,2}\alpha_1 z_k}\right|, & \theta_0 = 0 \\ C\left|e^{\lambda^0_{k,2}\alpha_2 z_k}\right|, & \theta_0 = \frac{5}{8}\pi, \frac{5}{4}\pi \end{cases}$$

3) $\lambda_k = \lambda_{k,3}$

$$\left|\psi(z_k, \lambda_{k,3})\right| \ge \begin{cases} C\left|e^{\lambda^0_{k,3}\alpha_2 z_k}\right|, & \theta_0 = 0, \frac{5}{8}\pi \\ C\left|e^{\lambda^0_{k,3}\alpha_1 z_k}\right|, & \theta_0 = \frac{11}{8}\pi \end{cases}$$

Let's T_{z_0} ($z_0 = |z_0|e^{\theta_0 i}$) be polygon with vertex:

1) if $\theta_0 \in (0, \frac{5}{8}\pi)$

$$|z_0|\cos(\theta_0 - \frac{\pi}{4})\cos^{-1}\frac{\pi}{4},\ |z_0|e^{\theta_0 i},\ |z_0|\cos(\theta_0 - \frac{9}{8}\pi)\cos^{-1}\frac{\pi}{4}e^{\frac{5}{8}\pi i},\ |z_0|e^{\left(\frac{5}{4}\pi - \theta_0\right)i},$$

$$|z_0|e^{\left(\frac{5}{4}\pi + \theta_0\right)i},\ |z_0|\cos(\theta_0 - \frac{3}{8}\pi)\cos^{-1}\frac{\pi}{4}e^{\frac{11}{8}\pi i},\ |z_0|e^{-\theta_0 i};$$

2) if $\theta_0 \in (\frac{5}{8}\pi, \frac{3}{4}\pi)$

$$|z_0|\cos(\theta_0 + \pi)\cos^{-1}\frac{\pi}{4},\ |z_0|\cos(\theta_0 - \frac{3}{8}\pi)\cos^{-1}(\theta_0 - \frac{5}{8}\pi)e^{\left(\theta_0 - \frac{3}{8}\pi\right)i},$$

$$|z_0|\cos(\theta_0 - \frac{3}{8}\pi)\cos^{-1}(\theta_0 - \frac{\pi}{2})e^{(\theta_0-\frac{\pi}{8})i},\ |z_0|e^{\theta_0 i},\ |z_0|\cos(\theta_0 - \frac{3}{8}\pi)\cos^{-1}\frac{\pi}{8}e^{(\frac{\pi}{8})},$$

$$|z_0|\cos(\theta_0 - \frac{3}{8}\pi)\cos^{-1}\frac{\pi}{4}e^{\frac{11}{8}\pi i},\ |z_0|\cos(\theta_0 - \frac{3}{8}\pi)\cos^{-1}(\theta_0 - \frac{5}{8}\pi)e^{-\left(\theta_0-\frac{3}{8}\pi\right)i};$$

3) if $\theta_0 \in (\frac{3}{4}\pi, \frac{5}{4}\pi)$

$$|z_0|\cos(\theta_0 + \pi)\cos^{-1}\frac{\pi}{4},\ |z_0|e^{\left(\frac{5}{4}\pi-\theta_0\right)i},\ |z_0|\cos^{-1}\frac{\pi}{4}\cos(\theta_0 + \frac{9}{8}\pi)e^{\frac{5}{8}\pi i}$$

$$|z_0|e^{\theta_0 i},\ |z_0|\cos^{-1}\frac{\pi}{4}\cos(\theta_0 - \pi)\cos(\theta_0 - \frac{9}{8}\pi)e^{\frac{11}{8}\pi i},\ |z_0|e^{\left(\theta_0-\frac{5}{4}\pi\right)i};$$

4) if $\theta_0 \in (\frac{5}{4}\pi, \frac{11}{8}\pi)$

$$|z_0|\cos(\theta_0 + \pi)\cos^{-1}\frac{\pi}{4},\ |z_0|\cos(\theta_0 + \frac{3}{8}\pi)\cos^{-1}(\theta_0 - \frac{5}{4}\pi)e^{(\theta_0-\pi)i},$$

$$|z_0|\cos(\theta_0 + \frac{3}{8}\pi)\cos^{-1}\frac{\pi}{4}e^{\frac{5}{8}\pi i},\ |z_0|\cos(\theta_0 + \frac{3}{8}\pi)\cos^{-1}(\theta_0 - \frac{3}{4}\pi)e^{\left(\theta_0-\frac{3}{8}\pi\right)i},$$

$$|z_0|\cos(\theta_0 + \frac{3}{8}\pi)\cos^{-1}(\theta_0 - \frac{3}{4}\pi)e^{\left(\theta_0-\frac{\pi}{4}\right)i},\ |z_0|\cos(\theta_0 + \frac{3}{8}\pi)\cos^{-1}\frac{\pi}{4}e^{\frac{11}{8}\pi i},$$

$$|z_0|\cos(\theta_0 + \frac{3}{8}\pi)\cos^{-1}(\theta_0 - \frac{5}{4}\pi)e^{(\pi-\theta_0)i};$$

5) if $\theta_0 \in (\frac{11}{8}\pi, 2\pi)$

$$|z_0|\cos(\theta_0 + \frac{\pi}{4})\cos^{-1}\frac{\pi}{4},\ |z_0|e^{(2\pi-\theta_0)i},\ |z_0|\cos^{-1}\frac{\pi}{4}\cos(\theta_0 + \frac{3}{8}\pi)e^{\frac{5}{8}\pi i},$$

$$|z_0|e^{\left(\theta_0-\frac{3}{4}\pi\right)i},\ |z_0|e^{-\left(\theta_0+\frac{\pi}{4}\right)i},\ |z_0|\cos(\theta_0 + \frac{3}{8}\pi)\cos^{-1}\frac{\pi}{4}e^{\frac{11}{8}\pi i},\ |z_0|e^{\theta_0 i};$$

6) if $\theta_0 = 0$

$$|z_0|,\ |z_0|\cos\frac{3}{8}\pi\cos^{-1}\frac{\pi}{4}e^{\frac{5}{8}\pi i},\ |z_0|\cos\frac{3}{8}\pi\cos^{-1}\frac{\pi}{8}e^{\pi i},\ |z_0|\cos\frac{3}{8}\pi\cos^{-1}\frac{\pi}{4}e^{\frac{11}{8}\pi i};$$

7) if $\theta_0 = \frac{5}{8}\pi$

$$|z_0|\cos\frac{3}{8}\pi\cos^{-1}\frac{\pi}{4},\ |z_0|e^{\frac{5}{8}\pi i},\ |z_0|e^{\frac{11}{8}\pi i};$$

8) if $\theta_0 = \frac{3}{4}\pi$

$|z_0|,\ |z_0|e^{\frac{3}{4}\pi i},\ |z_0|\cos\frac{3}{8}\pi\cos^{-1}\frac{\pi}{4}e^{\frac{11}{8}\pi i};$

9) if $\theta_0 = \frac{5}{4}\pi$

$|z_0|,\ |z_0|\cos\frac{3}{8}\pi\cos^{-1}\frac{\pi}{4}e^{\frac{5}{8}\pi i},\ |z_0|e^{\frac{5}{4}\pi i};$

10) if $\theta_0 = \frac{11}{8}\pi$

$|z_0|\cos\frac{3}{8}\pi\cos^{-1}\frac{\pi}{4},\ |z_0|e^{\frac{5}{8}\pi i},\ |z_0|e^{\frac{11}{8}\pi i}$

By the given lemmas and the results from the work [4] the following theorem is proved

Theorem 2. *If*

$$\sum_{k=1}^{\infty} a_k \lambda_k^s \psi(x, \lambda_k) \tag{4}$$

$$(s = 3p + \nu,\ p = 0, 1, 2, \ldots,\ \nu = 0, 1, 2)$$

from the series one of them is uniformly convergent in $z_0 = |z_0|e^{\theta_0 i}$ *point complex plane, where*

a) $\theta_0 \in (\gamma_j, \delta_j),\ 1 \le j \le 5,\ \gamma_1 = 0,$

$$\gamma_{j+1} = \delta_j = \frac{j + 4 + 3(\delta_{j,3} + \delta_{j,4})}{8}\pi,\ \delta_5 = 2\pi$$

b) $\theta_0 = 0,\ \gamma_{j+1},\ 2\pi \quad (1 \le j \le 4,\ |z_0| < 1).$

Then

1) (4) series are absolutely and uniformly convergent in any closed domain that wholly belongs to T_{z_0} and consequently in domain T_{z_0} it represents analytical function;

2) if f_ν $(\nu = 0, 1, 2)$ are the sums of the first three series then the sum result of every following three series is defined by the formula

$$f_s = f_{s-3}^{(3)} - a f_{s-2}^{(2)} + a f_{s-1}^{(1)} \quad (s = 3p + \nu,\ p = 1, 2, \ldots)$$

3) f_s functions satisfy boundary conditions to zero.

Georgian Subtropical Institute

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MATHEMATICS

N.Inasaridze

Non-Abelian Tensor Products of Finite Groups with Non-Compatible Actions

Presented by Corr. Member of the Academy H.Inasaridze, February 20, 1996.

ABSTRACT. **Some sufficient conditions for finiteness of a generalized non-abelian tensor product of groups are established.**

The non-abelian tensor product of groups was introduced by Brown and Loday [1-3].

It was defined for any A and B groups which act on themselves by conjugation and each of them acts on the other, so that the following compatibility conditions hold

$$^{(ab)}a'={}^{aba^{-1}}a',\quad {}^{(ba)}b'={}^{bab^{-1}}b', \tag{1}$$

for all $a,a'\in A$ and $b,b'\in B$ and $aba^{-1},bab^{-1}\in A*B$ (free product of groups).Conditions (1) are very important for the investigation of the tensor product of Brown and Loday. It is known from G.Ellis's paper [4] the finiteness of this tensor product and (1) is important to prove this fact too.

The non-abelian tensor product of groups introduced by Brown and Loday was not available for constructing of the non-abelian homology of groups [5]. Therefore a new tensor product was introduced [5] without conditions (1) and with extra condititions. Just this tensor product will be considered in this paper.

Definition 1. *Let A and B be arbitrary groups which act on themselves by conjugation and each of them acts on the other.The non-abelian tensor product $A\otimes B$ of the groups A and B is the group generated by the symbols $a\otimes b$ and defined by the relations*

$$aa'\otimes b=({}^{a}a'\otimes {}^{a}b)(a\otimes b)$$

$$a\otimes bb'=(a\otimes b)\,({}^{b}a\otimes {}^{b}b')$$

$$(a\otimes b)\,(a'\otimes b')=({}^{[a,b]}a'\otimes {}^{[a,b]}b')(a\otimes b)$$
$$(a'\otimes b')\,(a\otimes b)=(a\otimes b)\,({}^{[b,a]}a'\otimes {}^{[b,a]}b')$$

for all $a,a'\in A$ and $b,b'\in B$, where $[a,b]=aba^{-1}b^{-1}\in A*B$.

Remark 2. *If the actions of the pair A,B of groups satisfy the compatibility conditions (1) then we obtain the tensor product of Brown and Loday.*

In this paper we concern the finiteness of this tensor product of groups.In general it is an open problem, but we give some sufficent conditions of finiteness.

Definition 3 . *Let A and B be arbitrary groups which act on themselves by conjugation and each of them acts on the other. Then Com-subgroup of A with B denoted by Com $A(B)$ is the normal subgroup of A generated by the elements ${}^{(ab)}a'\,{}^{aba^{-1}}a'^{-1}$ where $a,a'\in A$, $b'\in B$.*

Note that if A and B act on each other compatibly then *Com* $A(B)$ and *ComB*(A) are trivial groups.

Definition 4. *Let A and B be groups which act on each other. Under these actions A acts on B perfectly if the action of A on B induces the action of A on Com. $B(A)$ and*

A,B are groups with perfect actions if under these actions A acts on B perfectly and B acts on A perfectly.

Definition 5. *Let (A,B) be an arbitrary pair of groups which acts on each other. Then Com-pairs of (A,B) are the pairs $(ComA(B),B)$ and $(A, Com\,B(A))$ of groups.*

Let (A,B) be an arbitrary pair of groups. On the zero stage take the pair (A,B) of groups. If (A,B) are with perfect actions then we can go over to the first stage and take its *Com*-pairs with induced actions. If the obtained pairs are with perfect actions then we can go over to the second stage and take their *Com*-pairs with induced actions if not,this process stop at the first stage and so on.

Definition 6. *The family of these obtained pairs will be called the compatibility resolution of the pair (A,B).*

Definition 7. *Let A and B be groups acting on each other. It should be said that A and B act on each other half compatibly if for every pair of groups (C,D) of the compatibility resolution of (A,B) the actions are perfect and the following conditions hold: ${}^{h}dd^{-1} \in ComD(C)$, ${}^{g}cc^{-1} \in ComC(D)$ for each $c \in C$, $d \in D$, $h \in ComC(D)$, $g \in ComD(C)$ and at some n-th stage, $n \geq 0$, of the compatibility resolution of (A, B) every pair (A_n, B_n) is a pair of groups with compatible actions.*

Clearly if A and B are groups which act on each other compatibly (i.e. conditions (1) hold) then they act on each other half compatibly.

Theorem 8. *Let A and B be finite groups acting on each other half compatibly, then $A \otimes B$ is finite.*

Theorem 9. *Let A and B be finite groups. Let A acts on B and B acts on A trivially.If $B^{(n)}$ is abelian for some $n \geq 1$, then $A \otimes B$ is finite,where $B^{(n)} = [B^{(n-1)}, B^{(n-1)}]$.*

Example 10. *Let A and B be finite groups. Let $[A,A]$ be abelian and $[A,A] \not\subset Z(A)$. Then the actions of $A \times B$ on A by conjugation and of A on $A \times B$ trivially do not satisfy the compatibility conditions (1).*
From Theorem 9 $(A \times B) \otimes A$ is finite.

Now it will be shown the existence of such finite group A. Suppose $M = Z_p^{+}$ (additive group) and $N = Z_p^{\times} \setminus \{0\}$ (multiplicate group) for any simple $p > 2$. Assume that N acts on M by multiplication of Z_p i.e. ${}^{[n]}[m] = [nm]$ for all $[n] \neq 0$, $[m] \in Z_p$. Let consider $M \rtimes N$ (semi-direct product of M and N). Then the commutant $[M \rtimes N, M \rtimes N] = M$ and therefore it is abelian. It can be shown that $[M \rtimes N, M \rtimes N] \not\subset Z(M \rtimes N)$.

Thus $M \rtimes N$ is an example of the above mentioned finite group A and therefore $\{(M \rtimes N) \times B\} \otimes (M \rtimes N)$ is finite for any finite group B.

Definition 11. *Let A and B be groups and A acts on B. Then $[A,B]$ is a normal subgroup of B generated by elements ${}^{a}bb^{-1}$ for all $a \in A$, $b \in B$, and we can define*

$$[A,B]^{n} = [[A,B]^{n-1}],\ n > 1,$$

since the action of A on B induces the action of A on $[A,B]$.

Theorem 12. *Let A and B be finite groups. A acts on B and B acts on A trivially. If $[A,B]^{n}$ is abelian for some $n \geq 1$,then $A \otimes B$ is finite.*

Note that Example 10 is available as example for Theorem 12.

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A.Razmadze Mathematical Institute
Georgian Academy of Sciences

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MATHEMATICS

M.Tchumburidze

About Matrices of Singular and Fundamental Solutions for Equations of the Generalised Couple - Stress Thermoelasticity Plane Theory

Presented by Academician T.Burchuladze, March 26, 1996

ABSTRACT. **The matrices of fundamental and other singular solutions are constructed explicit for the system of stationary equations of two - dimensional generalised couple - stress thermoelasticity theory and some of their properties are indicated.**

Basic dynamic homogeneous system of partial differential equations of generalized Green -- Lindsay couple - stress thermoelasticity on the plane for the homogeneous, izotropyc elastic media with the centre of symmetry has the form [1, 4, 5]:

$$L\left(\frac{\partial}{\partial x}, \frac{\partial}{\partial t}\right) v(x, t) = 0, \tag{1}$$

where,

$$L\left(\frac{\partial}{\partial x}, \frac{\partial}{\partial t}\right) = \left\| L_{j,k}\left(\frac{\partial}{\partial x}, \frac{\partial}{\partial t}\right)\right\|_{4\times 4},$$

$$L_{jk} = \delta_{jk}\left[(m+\alpha)\Delta - \rho\frac{\partial^2}{\partial t^2}\right], + (\lambda + \mu - \alpha)\frac{\partial^2}{\partial x_j\,\partial x_k}, j, k = 1,2$$

$$L_{jk} = \delta_{jk}\left[(m+\alpha)\Delta - 4\alpha - I\frac{\partial^2}{\partial t^2}\right], j = 3,\ k = 3,4$$

$$L_{jk} = -\,2\alpha\sum_{p=1}^{2}\varepsilon_{jkp}\frac{\partial}{\partial x_p}, \quad \begin{matrix} j = 1,2,\ k = 3 \\ k = 1,2,\ j = 3 \end{matrix}$$

$$L_{j4} = -\,\gamma_\tau\frac{\partial}{\partial x_j}, \quad j = 1,2$$

$$L_{4k} = -\,\eta\frac{\partial}{\partial t}\frac{\partial}{\partial x_k}, \quad k = 1,2$$

$$L_{43} = L_{34} = 0,$$

$$L_{44} = \Delta - \frac{1}{\chi_\tau}\frac{\partial}{\partial t}, \gamma_\tau = \gamma\left(1 + \tau_1\frac{\partial}{\partial t}\right), \frac{1}{\chi_\tau} = \frac{1}{\chi}\left(1 + \tau_0\frac{\partial}{\partial t}\right),$$

where $V = (v_1, v_2, v_3, v_4)^T = (v, v_3, v_4)^T$, $v = (v_1, v_2)$ is the displacement vector, v_3 - characteristic of the rotation and v_4 is the temperature variation, symbol T - transpose operator, $x = (x_1, x_2)$ is the point of the twodimensional Euclidean space R^2 and t is the time, Δ is the two dimensional Laplacian operator, δ_{kj} - symbol of Kronikery, ε_{jkp} - symbol of Levi-Chivita, $\gamma, \chi, \eta, \rho, \lambda, \mu, \alpha, \nu, \beta, I$ - constant which satisfies the following conditions [2]: $\mu > 0,\ I > 0,\ \rho > 0,\ \alpha > 0,\ \nu > 0,\ \beta > 0,\ \chi > 0,\ \lambda + \mu > 0,\ \frac{\gamma}{\nu} > 0,\ \tau_0,\ \tau_1$ - the constants of relaxations [2]: $\tau_1 \geq \tau_0 \geq 0$.

In case of harmonic oscillations and pseudooscillations the (1) system to the $V(x,\ \omega) = (u_1,\ u_2,\ u_3,\ u_4)^T = (U,\ u_4)^T,\ U = (u_1,\ u_2,\ u_3)$ has come to the following form [2, 4]:

$$L\left(\frac{\partial}{\partial x}, -i\omega\right)v(x,\ \omega) = 0, \tag{2}$$

where ω is real or complex parameter respectively.

Let us search the matrix of fundamental solutions:

$\Phi(x,\ \omega) = \|\Phi_{jk}\|_{4\times4}$

with the following form:

$$\Phi(x,\ \omega) = \hat{L}\left(\frac{\partial}{\partial x}, -i\omega\right)\varphi(x,\ \omega), \tag{3}$$

where $\hat{L}\left(\frac{\partial}{\partial x}, -i\omega\right)$ - is the connecting matrix of $L\left(\frac{\partial}{\partial x}, -i\omega\right)$:

$$\hat{L}\left(\frac{\partial}{\partial x}, -i\omega\right)L\left(\frac{\partial}{\partial x}, -i\omega\right) = L\left(\frac{\partial}{\partial x}, -i\omega\right)\hat{L}\left(\frac{\partial}{\partial x}, -i\omega\right) = I\det L\left(\frac{\partial}{\partial x}, -i\omega\right) \tag{4}$$

I - is the unit matrix of dimension 4×4, $\varphi(x,\ \omega)$ - is researching scalar function.

As it is known $\hat{L}\left(\frac{\partial}{\partial x}, -i\omega\right) = \left\|\hat{L}\left(\frac{\partial}{\partial x}, -i\omega\right)\right\|_{4\times4}$, $\hat{L}\left(\frac{\partial}{\partial x}, -i\omega\right)$ - algebrical addition of $L_{jk}\left(\frac{\partial}{\partial x}, -i\omega\right)$ element on the $L\left(\frac{\partial}{\partial x}, -i\omega\right)$ matrix. With (3) and (4) force we get differential equations of the eighth *row* according to the $\varphi(x,\ \omega)$:

$$\det L\left(\frac{\partial}{\partial x}, -i\omega\right)\varphi(x,\ \omega) = (\lambda + 2\mu)(\mu + \alpha)(\nu + \beta)\prod_{k=1}^{4}\left(\Delta + \sigma_k^2\right)\varphi(x,\ \omega) = 0, \tag{5}$$

where $\sigma_k^2,\ k = 1,4$ characteristic parametres, which are explained from the following equations:

$$\begin{aligned} \sigma_1^2 + \sigma_2^2 &= \rho_0\omega^2(1+\tau_2) + \frac{i\omega}{æ}(1+\varepsilon) \\ \sigma_1^2 \cdot \sigma_2^2 &= \rho_0 \frac{i\omega}{æ}\omega^2 + \omega^4\tau_3 \\ \sigma_3^2 + \sigma_4^2 &= \frac{\rho\omega^2}{\mu+\alpha} + \frac{I\omega^2 - 4\alpha}{\nu+\beta} + \frac{4\alpha^2}{(\mu+\alpha)(\nu+\beta)} \\ \sigma_3^2 \cdot \sigma_4^2 &= \frac{\rho\omega^2}{\mu+\alpha} \cdot \frac{I\omega^2 - 4\alpha}{\nu+\beta} \end{aligned} \tag{6}$$

Here, $\rho_0 = \dfrac{\rho}{\lambda+2\mu}$, $\tau_2 = \dfrac{\gamma\eta\tau_1}{\rho} + \dfrac{\lambda+2\mu}{\rho\chi}\tau_0 = \dfrac{1}{\rho_0\chi}(\varepsilon\tau_1+\tau_0)$, $\tau_3 = \dfrac{\rho_0\tau_0}{\chi}$, $\varepsilon = \dfrac{\chi\gamma\eta}{\lambda+2\mu}$.

According to (5) we have:

$$\varphi(\chi,\omega) = \sum_{k=1}^{4} a_k \cdot H_0^{(1)}(\sigma_k|x|), \tag{7}$$

where $H_0^{(1)}(\sigma_k|x|)$ - are Hunkel functions (the first kind the zero row), $|x| = \sqrt{x_1^2+x_2^2}$, a_k are constants. As it is known in the area close to zero we have Hunkel function with the following form [1]:

$$\frac{\Pi}{2i}H_0^{(1)}(\sigma_k|x|) = \ln|x| - \frac{\sigma_k^2}{2^2}|x|^2\ln|x| + \frac{\sigma_k^4}{2^6}|x|^4\ln|x| - \frac{\sigma_k^6}{2^8\cdot 3^2}|x|^6\ln|x| +$$
$$+ \text{const} + O|x|^8\ln|x|.$$

Let us, chose a_k, $k=\overline{1,4}$ constants so that 6th rows $\varphi(x,\omega)$ partial production close to zero area have speciality of $\ln|x|$ form. For this consider

$$a_k = \frac{i}{2\Pi(\lambda+2\mu)(\mu+\alpha)(\nu+\beta)(\sigma_{k+1}^2-\sigma_k^2)(\sigma_{k+2}^2-\sigma_k^2)(\sigma_{k+3}^2-\sigma_k^2)}, \quad k=\overline{1,4} \tag{8}$$

where, $\sigma_5=\sigma_1$, $\sigma_6=\sigma_2$, $\sigma_7=\sigma_3$.

If we check - up we get

$$\sum_{k=1}^{4}a_k = \sum_{k=1}^{4}a_k\sigma_k^2 = \sum_{k=1}^{4}a_k\sigma_k^4 = 0,\ \sum_{k=1}^{4}a_k\sigma_k^6 = \frac{i}{2\Pi(\lambda+2\mu)(\mu+\alpha)(\nu+\beta)},$$

$$\Phi_{pj}(x,\omega) = \sum_{k=1}^{4}\left\{\delta_{p\gamma}A_k + (-1)^{p+j}B_k\frac{\partial^2}{\partial x_{p+1}\partial x_{j+1}} + 4\alpha^2 C_k\frac{\partial^2}{\partial x_p\partial x_j}\right\}H_0^{(1)}(\sigma_k|x|),$$
$$x_3=x_1,\quad p,j=\overline{1,2}$$

$$\Phi_{pj}(x,\omega) = \sum_{k=3}^{4}\left\{N_K\sum_{n=1}^{2}\varepsilon_{pjn}\frac{\partial}{\partial x_n}\right\}H_0^{(1)}(\sigma_k|x|),\quad \begin{matrix} p=\overline{1,2}, & j=3 \\ j=\overline{1,2}, & p=3\end{matrix}$$

$$\Phi_{uj}(x,\omega) = -i\omega\eta\frac{\partial}{\partial x_j}\sum_{k=1}^{2}E_kH_0^{(1)}(\sigma_k|x|),\qquad j=\overline{1,2}$$

$$\Phi_{p4}(x,\omega) = \gamma(1-i\omega\tau_1)\frac{\partial}{\partial x_p}\sum_{k=1}^{2}E_kH_0^{(1)}(\sigma_k|x|),\quad p=\overline{1,2},$$

$$\Phi_{33}(x,\omega) = \sum_{k=3}^{4}D_kH_0^{(1)}(\sigma_k|x|),$$

$$\Phi_{44}(x,\omega) = \sum_{k=1}^{2}F_kH_0^{(1)}(\sigma_k|x|) + 8\alpha^2\frac{\partial^4}{\partial x_1^2\partial x_2^2}\sum_{k=1}^{4}a_kH_0^{(1)}(\sigma_k|x|),$$

where

$$A_k=\frac{a_k(\lambda_3^2-\sigma_k^2)(\lambda_4^2-\sigma_k^2)(\lambda_5^2-\sigma_k^2)}{(\mu+\alpha)(\nu+\beta)},\quad k=\overline{1,4},$$

$$B_k=\frac{a_k(\lambda_4^2-\sigma_k^2)(\lambda_6^2-\sigma_k^2)}{(\lambda+\mu-\alpha)(\nu+\beta)},\quad k=\overline{1,4}$$

$$C_k=a_k(\lambda_5^2-\sigma_k^2),\quad k=\overline{1,4},\quad N_k=\frac{a_k(\sigma_1^2-\sigma_k^2)(\sigma_2^2-\sigma_k^2)}{2\alpha(\lambda+2\mu)},\quad k=\overline{3,4}$$

$$D_k=\frac{a_k(\sigma_1^2-\sigma_k^2)(\sigma_2^2-\sigma_k^2)(\lambda_3^2-\sigma_k^2)}{(\mu+\alpha)(\lambda+2\mu)},\quad k=\overline{3,4}$$

$$E_k=\frac{a_k(\sigma_3^2-\sigma_k^2)(\sigma_4^2-\sigma_k^2)}{(\nu+\beta)(\mu+\alpha)},\quad k=\overline{1,2}$$

$$F_k=\frac{a_k(\sigma_3^2-\sigma_k^2)(\sigma_4^2-\sigma_k^2)(\lambda_7^2-\sigma_k^2)}{(\mu+\alpha)(\nu+\beta)(\lambda+2\mu)},\quad k=\overline{1,2}$$

It is easy to check - up, that

$$\sum_{k=1}^{4}B_k=\sum_{k=1}^{4}C_k=\sum_{k=3}^{4}N_k=\sum_{k=1}^{2}E_k=0,$$

where

$$\lambda_3^2=\frac{\rho\omega^2}{\mu+\alpha},\quad \lambda_4^2=\frac{I\omega^2-4\alpha}{\nu+\beta},\quad \lambda_6^2=\frac{\omega^2}{æ}(\tau_0+\tau_1\varepsilon_1)+\frac{i\omega}{æ}(1+\varepsilon_1),\quad \lambda_7^2=\rho_0\omega^2,$$

$$\varepsilon_1=\frac{\gamma\eta æ}{(\lambda+\mu-\alpha)}.$$

In the same way we can construct $\Phi(x,0)=\Phi(x)$ the matrix of fundamental solutions for static equations ($\omega=0$). In this case from (6) we have $\sigma_1=\sigma_2=\sigma_3=0$, $\sigma_4^2=-\dfrac{4\alpha\mu}{(\mu+\alpha)(\nu+\beta)}$.

The following notation is introduced: $\delta_4=\left[\dfrac{4\alpha\mu}{(\mu+\alpha)(\nu+\beta)}\right]^{1/2}$.

Taking this into account and carrying out some simple calculations, we find:

$$\varphi(x)=-\frac{\pi i}{2\delta_4^6}H_0^{(1)}(i\delta_4|x|)-\frac{1}{\delta_4^6}\ln|x|-\frac{1}{2^2\delta_4^4}|x|^2\ln|x|-$$

$$-\frac{1}{2^6\delta_4^2}|x|^4\ln|x|\equiv\frac{|x|^6\ln|x|}{2^8 3^2\pi(\lambda+2\mu)(\mu+\alpha)(\nu+\beta)}+O(|x|^8\ln|x|).$$

The elements of the matrix $\Phi(x)$ are equation:

$$\Phi_{kj}(x)=\frac{i\delta_{kj}\pi}{2\mu}(\mu+\alpha)^2(\nu+\beta)\left\{\frac{1}{\mu+\alpha}H_0^{(1)}(i\delta_4|x|)-\frac{2i}{\pi}\ln|x|\right\}+$$

$$+\frac{(-1)^{k+j+1}\pi i}{8\mu^2}(\lambda+\mu-\alpha)(\nu+\beta)^2(\mu+\alpha)\times$$

$$\times\frac{\partial^2}{\partial x_{k+1}\partial x_{j+1}}\left\{-H_0^{(1)}(i\delta_4(x))+\frac{2i}{\pi}\ln|x|+\frac{2i\mu}{\pi(\nu+\beta)}|x|^2\ln|x|\right\}-\frac{i(\mu+\alpha)^2(\nu+\beta)^2}{8\alpha\mu^2}\times$$

$$\times\frac{\partial}{\partial x_k\partial x_j}\left\{H_0^{(1)}(i\delta_4|x|-\frac{2i}{\pi}\ln|x|-\frac{i\alpha\mu}{\pi(\mu+\alpha)(\nu+\beta)}|x|^2\ln|x|\right\}+$$

$$+\delta_{kj}\frac{(13\mu+9\alpha+\lambda)(\nu+\beta)(\mu+\alpha)}{8\mu},\quad k,j=\overline{1,2},\ x_3=x_1$$

$$\Phi_{kj}(x)=-\frac{\pi i}{4\mu}(\lambda+2\mu)(\mu+\alpha)(\nu+\beta)\sum\varepsilon_{kjl}\ \partial\ \left\{H_0^{(1)}(i\delta_4|x|-\frac{2i}{\pi}\ln|x|\right\},\quad \begin{matrix}k=\overline{1,2},\ j=3\\ j=\overline{1,2},\ k=3\end{matrix}$$

$$\Phi_{33}(x)=-\frac{\pi i}{2}(\mu+\alpha)(\lambda+2\mu)H_0^{(1)}(i\delta_4|x|)$$

$$\Phi_{kj}(x)=\frac{\gamma(1-i\omega\tau_1)(\nu+\beta)(\mu+\alpha)}{2}(\ln|x|+1)x_k,\quad k=\overline{1,2},\ j=4$$

$$\Phi_{44}(x)=(\mu+\alpha)(\nu+\beta)(\lambda+2\mu)(\ln|x|+\frac{3}{2})+\frac{\alpha^2}{2^5 3^2}\frac{\partial^4}{\partial x_1^2\partial x_2^2}\left\{|x|^6\ln|x|+O(|x|^8\ln|x|)\right\}$$

$$\Phi_{41}=\Phi_{42}=\Phi_{34}=\Phi_{43}=0$$

From the rule of constructions of $\Phi(x,\omega)$ and $\Phi(x)$ we find

$$\varphi(x,\omega)-\varphi(x)=O(|x|^8\ln|x|)+\text{const} \qquad (10)$$

Near $|x|=0$ we have:

$$\Phi(x,\omega)-\Phi(x)=\hat{L}\left(\frac{\partial}{\partial x}\right)[\varphi(x,\omega)-\varphi(x)]+\left[\hat{L}\left(\frac{\partial}{\partial x},-i\omega\right)-\hat{L}\left(\frac{\partial}{\partial x}\right)\right]\varphi(x,\omega)=$$

$$=\hat{L}\left(\frac{\partial}{\partial x}\right)O(|x|^8\ln|x|)+\left[\hat{L}\left(\frac{\partial}{\partial x},-i\omega\right)-\hat{L}\left(\frac{\partial}{\partial x}\right)\right]|x|^6\ln|x|. \qquad (11)$$

Difference $\hat{L}\left(\frac{\partial}{\partial x},-i\omega\right)-\hat{L}\left(\frac{\partial}{\partial x}\right)$ contains not less than fourth row production according coordinats of Decart. That's why according to (11) we have:

$$\Phi(x,\omega)-\Phi(x)=O(|x|^2\ln|x|)$$

$$\frac{\partial^2}{\partial x_k\partial x_j}[\Phi(x,\omega)-\Phi(x)]=O(\ln|x|),\quad k,j=\overline{1,2}$$

Let $\widetilde{\Phi}(x,\omega)$ - the matrix of fundamental solutions of $\widetilde{L}\left(\frac{\partial}{\partial x},-i\omega\right)$ conjugate operation (the rule of Lagranje). It is easy to check - up equation:

$$\widetilde{\Phi}(x,\omega)=\Phi^T(-x,\omega)$$

With fundamental solutions it is possible to construct new singular solutions of (2) equation.

Let's discuss the differential operator:

$$HU=(TU-\gamma Nu_4);\quad RU=\left(HU,\frac{\partial u_4}{\partial n}\right)$$

where, $N=(n,0)^T$, $n=(n_1,n_2)$

TU - is the stress operator of the couple - stress elasticity [1, 2]

$$P^{(k)}U=(HU,-(\delta_{1k}+\delta_{2k})u_3+(\delta_{0k}+\delta_{3k})(\nu+\beta)\frac{\partial u_3}{\partial_n},\ -(\delta_{1k}+\delta_{3k})u_4+(\delta_{0k}+\delta_{2k})\frac{\partial u_4}{\partial n})^T,$$

$$Q^{(k)}U=(u,(\delta_{1k}+\delta_{2k})(\nu+\beta)\frac{\partial u_3}{\partial_n}+(\delta_{0k}+\delta_{3k})u_3,\quad (\delta_{1k}+\delta_{3k})\frac{\partial u_4}{\partial n}+(\delta_{0k}+\delta_{2k})u_4)^T$$

$$u=(u_1,u_2),\quad k=\overline{0,3},\quad P^{(0)}=R,\quad Q^{(0)}=\left\|\delta_{kj}\right\|_{4\times 4}$$

Let's construct matrices:

$$\left[\widetilde{R}\left(\frac{\partial}{\partial y},n\right)\widetilde{\Phi}(y-x,\omega)\right]^T=\left\|(\widetilde{R}\widetilde{\Phi})_{pj}\right\|^T_{4\times 4}$$

$$\left[\widetilde{P}^{(k)}\left(\frac{\partial}{\partial y},n\right)\widetilde{\Phi}(y-x,\omega)\right]^T=\left\|(\widetilde{P}^{(k)}\widetilde{\Phi})_{pj}\right\|^T_{4\times 4}$$

$$\left[\widetilde{Q}^{(k)}\left(\frac{\partial}{\partial y},n\right)\widetilde{\Phi}(y-x,\omega)\right]^T=\left\|(\widetilde{Q}^{(k)}\widetilde{\Phi})_{pj}\right\|^T_{4\times 4}$$

where, $\widetilde{R}$, $\widetilde{P}^{(k)}$, $\widetilde{Q}^{(k)}$ are conjugate operator the constructed matrices represent the solutions of the system (2). Let's call them basic singular solutions.

The constructed matrices have essential meaning in the theory of boundary value problems.

A.Razmadze Mathematical Institute
Georgian Academy of Sciences

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T.Aliashvili

Signature Method for Counting Point in Semi-Algebraic Subsets

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ABSTRACT. **The theorems on counting root numbers of real polynomial endomorphism in an arbitrary semi-algebraical set are given and several cases are considered. The estimate of the computational complexity of counting root numbers is obtained. The mentioned facts generalize early results obtained by the author. Particularly the numbers of variables and equations are unrestricted.**

Let $\varphi = (f, g)$: $\boldsymbol{R}^2 \to \boldsymbol{R}^2$ be a polynomial endomorphism of the type (m, n), i.e. $\deg f = m$ and $\deg g = n$. We call such endomorphism proper if its complexification $\varphi_C = \boldsymbol{C}^2 \to \boldsymbol{C}^2$ is proper in the usual sense, i.e. if the inverse image of the compact is compact.

From the $\boldsymbol{C}$- properness follows that there exist exactly $N = mn$ common complex roots $P_j = (\alpha_j, \beta_j), j = 0, \dots, N-1$. We assume that $\beta_i \neq \beta_j$, $i \neq j$ and such map is called y - convenient.

We use the classical Hermite - Jakoby signature method [1]. Namely, to a given map we assign the following quadratic form on the auxiliary space $\boldsymbol{R}^N$:

$$Q_h^\varphi(\bar{\xi}) = \sum_{j=0}^{N-1} h(\alpha_j, \beta_j)(\xi_0 + \xi_1\beta_j + \xi_2\beta_j^2 + \dots + \xi_{N-1}\beta_j^{N-1})^2 ,$$

where $h \in \boldsymbol{R}_2$ and $\boldsymbol{R}_2 = \boldsymbol{R}[x, y]$ is the ring of real polynomials in two variables. If $h \equiv 1$ we write $Q_1^\varphi \equiv Q^\varphi$ and call it principal "counting" form [2]. Denote by $\# Z_C(f, g, h)$ the number of complex roots of three functions $f, g, h \in \boldsymbol{R}_2$. We will describe how to compute $Z_C(f, g, h)$.

In the article [2] Q_h^φ always was a non - degenerated form. In the present paper it may be a degenerated form.

Theorem 1. *Let $\varphi: \boldsymbol{R}^2 \to \boldsymbol{R}^2$ be any convenient C - proper polynomial endomorphism and $h \in \boldsymbol{R}_2$ then*

$$\# Z_C(f, g, h) = N - rk\,(Q_h^\varphi).$$

We remark that $\# Z_C(f, g, h) \equiv \# Z_R(f, g, h) (\mathrm{mod}\ 2)$, so that if $N - rk\,(Q_h^\varphi)$ is odd then $\# Z_R(f, g, h) \neq \varnothing$.

Moreover, we can also estimate the real root number of a polynomial endomorphism in an arbitrary semi - algebraic set. It may be done with the help of formulae of the inclusion and elimination.

Theorem 2. *Let φ be of the same type as above and $h_1, h_2 \in \boldsymbol{R}_2$ such that $Z_C(f, g, h_i) = \varnothing$. $i = 1, 2$. Then*

$$\#\,(Z_R(f,g)\cap\{h_1>0,\,h_2>0\})=\frac{s(Q^{\varphi})+s(Q^{\varphi}_{h_1})+s(Q^{\varphi}_{h_2})+s(Q^{\varphi}_{h_1h_2})}{4}$$

where $s(Q^{\varphi})$, $s(Q^{\varphi}_{h_1})$ $s(Q^{\varphi}_{h_2})$ *and* $s(Q^{\varphi}_{h_1h_2})$ *are the signatures of the forms* Q^{φ}, $Q^{\varphi}_{h_1}$, $Q^{\varphi}_{h_{21}}$ *and* $Q^{\varphi}_{h_1h_2}$ *respectively.*

($Z_R(f,g)$ is the same that $\varphi_R^{-1}(0)$).

Remark 1. Particularly, if $h=J(f,g)$ is the Jacobian of φ then it is easily seen that $Z_R(f,g,J)$ coincides with the set of multiple roots of φ_C. Thus our formulae enable to calculate multiple root number for an arbitrary polynomial endomorphism.

Now we shall consider the case when there are three conditions of the inequality type $h_1, h_2, h_3\in R_2$. We are interested in $\#\,(Z_R(f,g)\cap\{h_1>0,\,h_2>0,\,h_3>0\})$.

In general case $\{h_1=0\}$ $\{h_2=0\}$ $\{h_3=0\}$ curves cut the plane at the eight pieces. Let $x_1, x_2, \ldots, x_8$ be the real root numbers in the corresponding domains, i.e. x_1 is the number of the roots which are in the domain $\{h_1,\ h_2,\ h_3\}>0$ $(+\ +\ +)$, $x_2-\{h_1>0,\,h_2>0$ and $h_3<0\}$ $(+\,+\,-)$, $x_3-(+\,-\,+)$, $x_4-(-\,+\,+)$, $x_5-(+\,-\,-)$, $x_6-(-\,+\,-)$, $x_7-(-\,-\,+)$, $x_8-(-\,-\,-)$. Let s, s_i, s_{ij} $(i,j=1,2,3;\ i\neq j)$ and s_{123} denote signatures of quadratic forms

Q^{φ}, $Q^{\varphi}_{h_i}$, $Q^{\varphi}_{h_ih_j}$ and $Q^{\varphi}_{h_1h_2h_3}$ respectively.

A simple calculation gives that

$$\#\,(Z_R(f,g)\cap\{h_1>0,\,h_2>0\ h_3>0\})=x_1=\frac{s+s_1+s_2+s_3+s_{12}+s_{13}+s_{23}+s_{123}}{8}.$$

It is possible to calculate the root number for every combination of h_1, h_2, h_3. For example,

$$x_5=\frac{s+s_1-s_2-s_3-s_{12}-s_{13}+s_{23}+s_{123}}{8}.$$

The system matrix has a special symmetry which enables us to use this method for arbitrary number of unequality condiditions. Thus we can solve the analogous problem for any semi - algebraic subsets of R^n.

Having in mind computations on personal computers it is interesting how much algebraic operations are needed for the estimation of the real root number of a polynomial endomorphism .

Using some results of the number theory we have obtained the following estimate of the computational complexity.

Proposition 1. *Let φ be a C - proper y - convenient polynomial endomorphysm of R^2 with the components having degrees not greater than d. Then the number of algebraic operations on the coefficients of components of the endomorphism φ which are needed for calculation of the signature Q^{φ} is less than the number*

$$2d^2(d+4)(2d)!-d(d+4)-\sum_{k=1}^{d+1}\ \sum_{i\geq\left[\frac{k+1}{2}\right]}^{d}(2i-k)!(i-1)+$$

$$+\sum_{p=1}^{2(d^2-1)} 6p\left(\left[\frac{e^{\pi\sqrt{\frac{2}{3}}\sqrt{p-\frac{1}{24}}}}{4\sqrt{3}\left(p-\frac{1}{24}\right)}\right]+1\right)+\sum_{k=1}^{d^2}\left(2k^2+8k+5\cdot\frac{(k-1)(k-2)}{2}\right)+2$$

It should be noted that this formula gives a sufficiently exact upper estimate.

It means that there exists such a choice of the coefficients that the above number may be attained.

It will be also interesting to compare this with the computational complexity of the algorithm announced by Z.Szafraniec in [3]. Unfortunately no details were published till now.

By means of the result described above and in [2] more general theorems are obtained. In this case we consider more polynomials.

Theorem 3. *If $\varphi = (f, g)$: $R^2 \to R^2$ C proper y - convenient polynomial endomorphysm and R_2 then*

$$\#(Z_R(f, g, h)) = s(Q^\varphi) - s(Q^\varphi_{(h^2)}).$$

This is the criterion of existence of common zeroes of three polynomials.

For arbitrary number of polynomials we have an analogous result.

Theorem 4. *Let φ be the same and $h_1, \ldots, h_l \in R_2$ then*

$$\#(Z_R(f, g, h_1, \ldots, h_l)) = s(Q^\varphi) - s(Q^\varphi_{(h^2)}),$$

where (h^2) denotes $\sum_{i=1}^{l} h_i^2$.

The more general case when we have $n + p$ functions of n variables is represented analogously to the preceeding case but the formula correspondingly acquires more contentional loading.

Obviously, everything here may be reduced to a calculation of Newton mixed sums [4], [5], which is a completely nonevident task.

Up-to-date discoveries concerning n - dimensional Grothendique residues [6] (cf. also [7]) enable us to perform this calculation in the general case. The precise method will be described in the next publication.

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Georgian Technical University

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MATHEMATICS

Corr. Member of the Academy N.Berikashvili

On the Criterion of the Secondary Cross Section

Presented June 13, 1996

ABSTRACT. **New formula for the secondary obstruction to Serre fibration is given.**

1. **Introduction.** Let $F \to E \to B$ be a Serre fibration with the base polyhedron and with the first obstructon zero. Hence if $\pi_i(F) = 0$, $i < q$, $i \neq p$ and $\pi_p = \pi_p(F)$, $\pi_q = \pi_q(F)$, then there exists a cross section s over q-skeleton and the so called secondary obstruction class $h^{q+1}(s^q) \in H^{q+1}(B, \pi_q)$ is defined which in contrast with the first obstruction class is not defined uniquely. The problem to decide whether the fibration has a cross section over $(q+1)$-skeleton of B in terms of secondary obstruction class and invariants of fibration was considered beginning Hopf [1] and in wide special cases was solved [1,2,3,4,5,6]. In [7] the author has constructed the "secondary obstruction functor" which provides alternative approach to the question and leads to an answer without exeption. The formulation of final answer does not use the notion of the mentioned functor and proceeds in familiar algebraic topology terminology. This formulation is given below.

2. **Statement of results.** Assume that we have a Serre fibration $F \to E \to B$ with fiber $\pi_i(F) = 0$, $i < q$, $i \neq p$. Let $\pi_p = \pi_p(F)$, $\pi_q = \pi_q(F)$. Let $K(G, n)$ denote the Eilenberg-MacLane complex. Consider the bicomplexes

$$C^*(B, C^*(K(\pi_p, p), \pi_q)) \tag{1}$$

and

$$C^*(B, C_*(K(\pi_p, p))). \tag{2}$$

They are regarded as cochain complexes of B with coefficient complexes

$$C^*(K(\pi_p, p), \pi_q) = Hom(C_*(K(\pi_p, p), \pi_q) \tag{3}$$

and

$$C_*(K(\pi_p, p)) \tag{4}$$

respectively. Of course (3) and (4) are paired by evaluation map to the group π_q:

$$C^*(K(\pi_p, p), \pi_q) \otimes C_*K(\pi_p, p)) \to \pi_q \tag{5}$$

and this pairing is differential.

The total degree in complex (2) is $p - q$ and we assume that in total complex n-cochains are given by $X^n = \prod_{p-q=n} C^p(B, C_q(K(\pi_p, p)))$. The filtration by first degree is complete in a sence of Eilenberg-Moore. The total complex is nontrivial in negative dimensions as well and our special attention is concerned especially to dimension 0. Of course

$$C^*(B, C_*(K(\pi, p))) = Hom(C_*(B), C_*(K(\pi, p))) \tag{6}$$

and hence a 0-dimensional cocycle of complex (2) is identified as a chain map

$$C_*(B) \to C_*(K(\pi, p)).$$

The complex (1) we can identify as the cell cochain complex of cartesian product of cell complexes B and $K(\pi_p, p)$:

$$C^*(B, C^*(K(\pi_p, p), \pi_q)) = C^*(B \times K(\pi_p, p), \pi_q)). \tag{7}$$

The pairing of coefficients (5) enables to define the pairing by $\cup$- product of total complexes (1) and (2) in $C^*(B, \pi_q)$:

$$C^*(B, C^*(K(\pi_p, p), \pi_q)) \otimes C^*(B, C_*(K(\pi_p, p))) \to C^*(B, \pi_q). \qquad (8)$$

This pairing defines the pairing of homology

$$H^*(C^*(B, C^*(K(\pi_p, p), \pi_q))) \otimes H^*(C^*(B, C_*(K(\pi_p, p)))) \to H^*(B, \pi_q)$$

i.e.

$$H^*(B\times K(\pi_p, p), \pi_q)) \otimes H^*(C^*(B, C_*(K(\pi_p, p)))) \to H^*(B, \pi_q). \qquad (9)$$

Especially any 0-dimensional cohomology class of complex (2) defines the homomorphism of groups $H^i(B\times K(\pi, p), \pi_q) \to H^i(B, \pi_q)$, $i = 0, 1, 2, 3, \dots$

In the complex $C^*(B\times K(\pi_p, p), \pi_q)$ we have two subcomplexes

$$0 \to C^*(K(\pi_p, p), \pi_q) \to C^*(B\times K(\pi_p, p), \pi_q). \qquad (10)$$

and

$$0 \to C^*(B, \pi_q) \to C^*(B\times K(\pi_p, p), \pi_q) \qquad (11)$$

These subcomplexes are direct summands and hence

$$C^*(B\times K(\pi_p, p), \pi_q) = C^*(K(\pi_p, p), \pi_q) + \overline{C}^*(B\times K(\pi_p, p), \pi_q) + C^*(B, \pi_q)$$

for suitable $\overline{C}^*(B\times K(\pi_p, p), \pi_q)$. (If B has only one 0-cell $*$ then this is obvious: $* \times K(\pi_p, p) \subset B\times K(\pi_p, p)) \supset B\times *$).

Let $s^q : B^q \to E$ be a cross section over q-skeleton. Then it defines an obstruction cocycle $c^{q+1}(s^q) \in C^{q+1}(B, \pi_q)$. With this analogy we can construct a $(q+1)$-cocycle on $B\times K(\pi_q, p)$: consider induced by projection $B\times K(\pi_p, p) \to B$ from fibration E a fibration $E' \to B\times K(\pi_p, p)$; let on the $(p-1)$-skeleton $[B\times K(\pi, p)]^{p-1} = [B\times *]^{p-1} = B^{p-1}$ the cross section $s_{B\times L}$ coincide with s^q; on the p-cell $\sigma^0 \times \tau^p$ let $s_{B\times L}$ represent τ^p as element of homotopy group π_p, $\tau^p \in \pi_p$; on the p-cell $\tau^p \times \tau^0$ let $s_{B\times L}$ coincide with s^q; constructed cross section extends from p-skeleton of $B\times K(\pi_p, p)$ to q-skeleton of $B\times K(\pi_p, p)$ and we can assume that it coincides over $(B \times *)^q$ with s^q. The obstruction cocycle $c^{q+1}(s^q{}_{B\times L}) \in C^{q+1}(B\times K(\pi_p, p)\pi_q)$ has the form:

$$c^{q+1}(s_{B\times K}) = m^{0,q+1} + m^{1,q} + m^{2,q-1} + m^{3,q-2} + \dots + m^{q-p+1,p} + \dots + 0 + \dots + 0 + m^{q+1,0},$$

where

$$m^{i,j} \in C^i(B, C^j(K(\pi_p, p)\pi_q))$$

and

$$m^{q+1,0} = c^{q+1}(s^q).$$

By above remark about cochain complex $C^*(B\times K(\pi_p, p)\pi_q)$ the above sum without $m^{q+1,0}$ is still cocycle. One sees that $m^{0,q+1}$ over every 0-simplex σ^0 is the k-invariant of fiber F. Without loss of generality we can assume that this cochain in $C^{q+1}(K(\pi_p, p), \pi_q)$ does not depend on σ^0 and hence $m^{0,q+1}$ is cocycle of $C^*(B\times K(\pi_p, p), \pi_q)$ too and lies in subcomplex (10) mentioned above. Thus we have

$$c^{q+1}(s_{B\times K}) = m^{0,q+1} + \overline{c}^{q+1}(s^q{}_{B\times K}) + c^{q+1}(s^q),$$

where $m^{0,q+1} \in Z^{q+1}(K(\pi_p, p), \pi_q)$ represents Postnikov invariant of fiber, $k^{q+1} = k^{q+1}(F) \in H(K(\pi_p, p), \pi_q)$, and

$$\overline{c}^{q+1}(s^q{}_{B\times K}) = m^{1,q} + m^{2,q-1} + m^{3,q-2} + \dots + m^{q-p+1,p} \in Z^{q+1}(B\times K(\pi_p, p), \pi_q).$$

Hence on homology level we have

$$h^{q+1}(s_{B\times K}) = k^{q+1}(F) + \overline{h}^{q+1}(s^q{}_{B\times K}) + h^{q+1}(s^q).$$

This class of course is the Postnikov invariant of the fibration E.

Lemma 1. *If $q < 2p$ then the homology class $\overline{h}^{q+1}(s^q{}_{B\times K})$ is invariant of fibration.*

Now for the complex $C^*(B, C_*(K(\pi_p, p)))$, in the case $B = K(\pi_p, p)$ there is a characteristic class $1+\chi^0$, $1+\chi^0 \in H(C^*(B. C_*(K(\pi, p))))$, of dimension 0, the homology class of the cocycle id: $C_*(K(\pi, p)) \to C_*(K(\pi_p, p))$-. id has in $C^*(K(\pi_p, p), C_*(K(\pi_p, p)))$ the form

$$f = f^{0,-0} + f^{1,-1} + f^{2,-2} + f^{3,-3} + \dots + f^{p,-p} + f^{p+1,-p-1} + \dots + f^{k,-k} + \dots$$

$$f = 1 + 0 + 0 + 0 + \dots + 0 + f^{p,-p} + f^{p+1,-p-1} + \dots + f^{k,-k} + \dots$$

$$f^{i,-i} \in C^i(K(\pi_p, p). C_i(K(\pi_p, p))).$$

Homology class $x^{0'}$ has the filtration p.

Let jet $x^p \in H^p(B, \pi_p)$. We consider it as a simplicial map x^p: $B \to K(\pi, p)$ and let $(x^p)^*(\chi^0)$ be image of class χ^0 in $H^0(C^*(B, C_*(K(\pi_p, p)))$.

Let now $\bar{s}_{B^q}$ be any other cross section over the q-skeleton of base and $h(\bar{s}_{B^q})^{q+1}$ its obstruction class. The principal result is

Theorem 1. *In above assumptions there is a cohomology class $x^p \in H^p(B, \pi_p)$ such that $h(s_{B^q})^{q+1} - h(\bar{s}_{B^q})^{q+1} = Op_{k^{q+1}}(x^p) + \overline{h}^{q+1}(s^q{}_{B\times K}) \circ (x^p)^*(\chi^0)$, where $\circ$ denotes the pairing (9) and Op_y denotes the cohomology operation defined by cohomology class $y \in H^*(K(\pi, p), \pi_q)$. For every $x^p \in H^p(B, \pi_p)$ there is a cross section $\bar{s}_{B^q}$ over the q-skeleton of base such that this equality holds.*

As immediate corollary one has the following criterion.

Theorem 2. *In above notations the cross section on the $(q+1)$-skeleton of base exists if and only if there is such a homology class $x^p \in H^p(B, \pi_p)$, that $Op_{k^{q+1}}(x^p) + \overline{h}^{q+1}(s^q{}_{B\times K}) \circ (x^p)^*(\chi^0) = h^{q+1}(s^q{}_B)$ where $\circ$ denotes the pairing (9) and Op_y denotes the cohomology operation defined by cohomology class $y \in H^*(K(\pi, p), \pi_q)$.*

A.Razmadze Mathematical Institute
Georgian Academy of Sciences

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MATHEMATICS

M.Usanetashvili

Conjugation Problem for a Second Order Hyperbolic System with Discontinuous Coefficients

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ABSTRACT. **The conjugation problem is considered for a second order degenerated hyperbolic system with discontinuous coefficients. The uniqueness and existence theorems are proved.**

Let us consider the hyperbolic system of linear differential equations

$$L(u)=\begin{cases} y^m A_1U^+_{xx}+2y^{\frac{m}{2}}B_1U^+_{xy}+C_1U^+_{yy}+a_1U^+_x+b_1U^+_y+C_1U^+=F_1, & x>0 \\ y^m A_2U^-_{xx}+2y^{\frac{m}{2}}B_2U^-_{xy}+C_2U^-_{yy}+a_2U^-_x+b_2U^-_y+C_2U^-=F_2, & x<0 \end{cases} \quad (1)$$

where A_i, B_i, C_i, a_i, b_i, c_i $(i=1,2)$ are given $(n\times n)$ real matrices, F_i $(i=1,2)$ are given functions and $U^{\pm}$ are unknown n-dimensional vectors, $m>0$, $n>1$.

We assume that A_i, B_i, C_i $(i=1,2)$ are constant matrices, $\det C_i\neq 0$ $(i=1,2)$ and polynomials

$$P_1(\lambda)=\det(A_1+2B_1\lambda+C_1\lambda^2),$$
$$P_2(\mu)=\det(A_2+2B_2\mu+C_2\mu^2),$$

have only the real different roots $\lambda_1, \lambda_2,\dots, \lambda_{2n}$ and $\mu_1, \mu_2,\dots, \mu_{2n}$ satisfying the following conditions

$$\lambda_1<\lambda_2<\dots<\lambda_n<0<\lambda_{n+1}<\dots<\lambda_{2n},$$
$$\mu_1<\mu_2<\dots<\mu_n<0<\mu_{n+1}<\dots<\mu_{2n} \quad (2)$$

In this case the system (1) is strictly hyperbolic and it has a parabolic degeneration when $y=0$. Under the given conditions the numbers $y^{\frac{m}{2}}\lambda_1,\dots, y^{\frac{m}{2}}\lambda_{2n}$ and $y^{\frac{m}{2}}\mu_1,\dots,y^{\frac{m}{2}}\mu_{2n}$ are roots of the characteristic polynomials of the system (1) $P_1(y,\lambda)=\det(y^mA_1+2y^{\frac{m}{2}}B_1\lambda+C_1\lambda^2)$ and $P_2(y,\mu)=\det(y^{\frac{m}{2}}A_2+2y^{\frac{m}{2}}B_2\mu+C_2\mu^2)$ respectively. In addition the characteristics of the system (1) passing through the point $P(x_0,y_0)$, $y_0>0$ satisfy the equations:

$$x+\frac{2\lambda_i}{m+2}y^{\frac{m+2}{2}}=x_0+\frac{2\lambda_i}{m+2}y^{\frac{m+2}{2}}y_0^{\frac{m+2}{2}}, \quad x_0>0, \quad i=1,2,\dots,2n,$$
$$x+\frac{2\mu_j}{m+2}y^{\frac{m+2}{2}}=x_0+\frac{2\mu_j}{m+2}y_0^{\frac{m+2}{2}}, \quad x_0<0, \quad j=1,2,\dots,2n.$$

Let D be a bounded domain in the upper half-plane $(y>0)$, and let it be surrounded by the following two characteristics of the system (1) passing through the origin $O(0,0)$

$$\gamma_1:\ x+\frac{2\lambda_n}{m+2}\,y^{\frac{m+2}{2}}\,y^{\frac{m+2}{2}}=0;\quad \gamma_2:\ x+\frac{2\mu_{n+1}}{m+2}\,y^{\frac{m+2}{2}}=0,$$

and by the two characteristics, passing through the point $O_1(0,y_0)$

$$\gamma_3:\ x+\frac{2\lambda_{2n}}{m+2}\,y^{\frac{m+2}{2}}=\frac{2\lambda_{2n}}{m+2}\,y_0^{\frac{m+2}{2}},\quad \gamma_4:\ x+\frac{2\mu_1}{m+2}\,y^{\frac{m+2}{2}}=\frac{2\mu_1}{m+2}\,y_0^{\frac{m+2}{2}},$$

here $y_0>0$ is an arbitrary fixed number. Let us denote by P_1 and P_2 the intersection points of γ_1 and γ_2, and of γ_3 and γ_4, respectively. Let $D^+\subset D$ be the domain surrounded by the curves γ_1 and γ_3 and by $x=0$, and $D^-\subset D$ – the domain surrounded by the γ_2 and γ_4 and $x=0$.

Further let us consider the following characteristic problem [1]: Find the regular solution of the system (1)

$$U(x,y)=\begin{cases}U^+(x,y), & (x,y)\in D^+,\\ U^-(x,y), & (x,y)\in D^-,\end{cases}$$

which satisfies the boundary conditions:

$$\left(y^{\frac{m}{2}}M_1\frac{\partial U^+}{\partial x}+N_1\frac{\partial U^+}{\partial y}+S_1U^+\right)\Big|_{op_1}=f_1, \tag{3}$$

$$\left(y^{\frac{m}{2}}M_2\frac{\partial U^-}{\partial x}+N_2\frac{\partial U^-}{\partial y}+S_2U^-\right)\Big|_{op_2}=f_2 \tag{4}$$

and the conjugation conditions on the segment OO_1:

$$U^+(0,y)-\Lambda_1U^-(0,y)=g_1(y),\quad 0\le y\le y_0, \tag{5}$$

$$U_x^+(0,y)-\Lambda_2U_x^-(0,y)=g_2(y),\quad 0\le y\le y_0 \tag{6}$$

here M_i, N_i, S_i, Λ_i $(i=1,2)$ are given real $(n\times n)$ matrices, Λ_i $(i=1,2)$ are constant matrices, f_i and g_i $(i=1,2)$ are given real n-dimensional vectors.

In the sequel we will provide that $a_1, b_1, c_1, F_1\in C^1(D^+)$, $a_2, b_2, c_2, F_2\in C^1(D^-)$, M_i, N_i, S_i, $f_i\in C^1(Op_i)$, $(i=1,2)$, $g_i\in C^1(00_1)$, $(i=1,2)$. Moreover, the inequalities

$$\sup_{D^+\backslash 0}\left\|y^{1-\frac{m}{2}}a_1\right\|<\infty,\quad \sup_{D^+\backslash 0}\left\|y^{1-\frac{m}{2}}a_{1x}\right\|<\infty,$$

$$\sup_{D^-\backslash 0}\left\|y^{1-\frac{m}{2}}a_2\right\|<\infty,\quad \sup_{D^-\backslash 0}\left\|y^{1-\frac{m}{2}}a_{2x}\right\|<\infty,$$

$$\sup_{D^+\backslash 0}\left\|y^{-\left(\alpha+\frac{m}{2}-1\right)}F_1\right\|<\infty,\quad \sup_{D^+\backslash 0}\left\|y^{-(\alpha-2)}F_{1x}\right\|<\infty$$

$$\sup_{D^-\backslash 0}\left\|y^{-\left(\alpha+\frac{m}{2}-1\right)}F_2\right\|<\infty,\quad \sup_{D^-\backslash 0}\left\|y^{-(\alpha-2)}F_{2x}\right\|<\infty, \tag{7}$$

$$f_i(0)=g_i(0)=0,\ (i=1,2),\ \alpha=\text{const}>0,$$

$$\sup_{Op_i\setminus 0}\left\| y^{-\left(\alpha+\frac{m}{2}\right)} f_i \right\| < \infty, \quad \sup_{Op_i\setminus 0}\left\| y^{-\left(\alpha+\frac{m}{2}\right)} f_i' \right\| < \infty,$$

$$\sup_{OO_1\setminus 0}\left\| y^{-\left(\alpha+\frac{m}{2}\right)} g_i \right\| < \infty, \quad \sup_{OO_1\setminus 0}\left\| y^{-\left(\alpha+\frac{m}{2}\right)} g_i' \right\| < \infty, \quad i=1,2$$

hold in D^+ and D^-.

The solution of the problem (1),(3)-(6) is sought in the class

$$\left\{ U^\pm \in C^2(D^\pm): U^\pm(0,0)=0, \quad \sup_{\overline{D}^+\setminus 0}\left\| y^{-\alpha} U_x^+ \right\| < \infty, \quad \sup_{\overline{D}^-\setminus 0}\left\| y^{-\alpha} U_x^- \right\| < \infty \right.$$

$$\sup_{\overline{D}^+\setminus 0}\left\| y^{-\left(\alpha+\frac{m}{2}\right)} U_y^+ \right\| < \infty, \quad \sup_{\overline{D}^-\setminus 0}\left\| y^{-\left(\alpha+\frac{m}{2}\right)} U_y^- \right\| < \infty \qquad (8)$$

Because of the fact that $\lambda_1, \lambda_2, \ldots, \lambda_{2n}$ and μ_1, μ_2, μ_{2n} are the simple roots of the polynomials $P_1(\lambda)$ and $P_2(\mu)$, the following equations hold

$$\dim\mathrm{Ker}(A_1+2B_1\lambda_i+C_1\lambda_i^2)=1, \quad \dim\mathrm{Ker}(A_2+2B_2\mu_j+C_2\mu_j^2)=1$$

$$1\le i,j\le 2n$$

Let us denote by ν_i and ν_j^* the vectors $\nu_i \in \mathrm{Ker}(A_1+2B_1\lambda_i+C_1\lambda_i^2)$, $\|\nu_i\|\neq 0$, $\nu_j^* \in \mathrm{Ker}(A_2+2B_2\mu_j+C_2\mu_j^2)$, $\|\nu_j^*\|\neq 0$, $i,j=1,\ldots,2n$ where $\|\cdot\|$ denotes the norm in R^n. Let us introduce the matrices

$$\Gamma_i(M_i,N_i), \quad i=1,2, \quad K=\begin{pmatrix} K_1 \\ K_2 \end{pmatrix}=(V_1, V_2), \; K^*=\begin{pmatrix} K_1^* \\ K_2^* \end{pmatrix}=(V_1^*, V_2^*),$$

$$K=\begin{pmatrix} \nu_1,\ldots,\nu_{2n} \\ \lambda_1\nu_1,\ldots,\lambda_{2n}\nu_{2n} \end{pmatrix}, \quad \widetilde{K}=\begin{pmatrix} y^{-\frac{m}{2}}\nu_1,\ldots,y^{-\frac{m}{2}}\nu_{2n} \\ \lambda_1\nu_1,\ldots,\lambda_{2n}\nu_{2n} \end{pmatrix},$$

$$K^*=\begin{pmatrix} \nu_1^*,\ldots,\nu_{2n}^* \\ \mu_1\nu_1^*,\ldots,\mu_{2n}\nu_{2n}^* \end{pmatrix}, \quad \widetilde{K}^*=\begin{pmatrix} y^{-\frac{m}{2}}\nu_1,\ldots,y^{-\frac{m}{2}}\nu_{2n} \\ \lambda_1\nu_1,\ldots,\lambda_{2n}\nu_{2n} \end{pmatrix},$$

$$\widetilde{K}_1=y^{-\frac{m}{2}}K_1, \quad \widetilde{K}_2=K_2, \quad \widetilde{K}_1^*=K_2^*, \quad G_1=\begin{pmatrix} O \\ E \end{pmatrix}, \quad G_2=\begin{pmatrix} E \\ O \end{pmatrix}.$$

Theorem. *Let the conditions*

$$\det(\Gamma_1\times V_1)(x,y)\neq 0, \qquad (x,y)\in OP_1,$$

$$\det(\Gamma_2\times V_2^*)(x,y)\neq 0, \qquad (x,y)\in OP_2,$$

$$\det\begin{pmatrix} \widetilde{K}_2G_1-\Lambda_1\widetilde{K}_2^*G_2 \\ \widetilde{K}_1G_1-\Lambda_2\widetilde{K}_1^*G_2 \end{pmatrix}(x,y)\neq 0, \; (x,y)\in(0,0_1).$$

be fulfilled.

Then there exists a positive number depending only on the coefficients of the system in case if $\alpha>\alpha_0$. The problems (1)- (6) have a unique solution in the class of vector-functions introduced above.

I.Javakhishvili Tbilisi State University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №1, 1996

MATHEMATICS

N.Skhirtladze

Some Properties of Belman's and Loo's Transformations

Presented by Academician L.Zhizhiashvili, May 23, 1996

ABSTRACT. **Some properties of Belman's and Loo's trasformations and invariant classes for this transformations are studied.**

I. Let f belong to class $C(I)$. It means that f is 2π-periodic and continuous function. Imply that $\delta\in]0, 2\pi[$ and

$$\Delta_n^k(f;x)=\sum_{i=0}^{k}(-1)^{k-i}\binom{k}{i}f(x+ih).$$

Expression [2], $\omega^{(k)}(\delta;f)_C=\sup_{|h|\le\delta}\left\|\Delta_h^k(f;.)\right\|_C$

is modulus of fluency of order $-k$

Imply that ω is modulus of continuity [2].

H^{ω} - defines the class of functions

$H^{\omega}=\{f;\omega(\delta;f)_C\le A(f)\omega(\delta)\}$ where $A(f)\in]0,\infty[$.

In future we need some conditions of N.K.Bari and S.B.Stechkin [3] on continuity:

Exactly $\int_0^{\delta}t^{-1}\omega(t)dt+\delta\int_{\delta}^{\pi}t^{-2}\omega(t)dt\le A\omega(\delta)$.

II. If $f\in L(T)$ and is even, then its Fourier series of function is

$$\sigma[f]=\frac{a_0}{2}+\sum_{n=1}^{\infty}a_n\cos nx. \qquad (1)$$

And if f is odd function

$$\sigma[f]=\sum_{n=1}^{\infty}b_n\sin nx. \qquad (2)$$

Imply that Fourier series of function f is given by (2) and for every natural number n the following value is considered

$$B_n=\sum_{k=n}^{\infty}\frac{b_k}{k} \qquad (3)$$

This expression is called transformation of Bellman. Loo [5] considered function - ψ. Imply that $f\in L\,ln^{+}L(T)$ and f is odd function. If

$$\psi(x)\equiv\psi(x;f)=\frac{1}{tg\,x/2}\int_0^{x}f(t)dt, \qquad (4)$$

where $\psi(0)=\psi(\pi)=0$, then ψ is called transformation of Loo.

If we consider the conjugative function $\overline{\psi}$, of function ψ [6], it is unknown whether ψ is integrable on T, or not.

Following theorem is correct

Theorem 1. *Let $f\in(ln^+L)^2(T)$, then $\overline{\psi}$ is integrable on T.*

If $b_k \geq 0$ ($k=1,2,3,\ldots$), then from (3) $(B_k-\frac{b_k}{2k})_{k\geq 1}$ sequence tends to -0 with monotony.

N.K.Bari and S.B.Stechkin's conditions can be written as follows:

$$\frac{1}{n}\sum_{k=1}^{n}\omega\left(\frac{1}{k}\right)\leq A\omega\left(\frac{1}{n}\right). \tag{5}$$

Theorem 2. *Let $\sigma[f]$ be given by (2) and $b_k \geq 0$, ($k = 1,2,3\ldots$).*

Imply that ω is modulus of continuity for which following condition (5) is correct. If

$$\sum_{k=n}^{\infty}\left(B_k-\frac{b_k}{2k}\right)=0\left(\omega\left(\frac{1}{n}\right)\right),\ then\ f\in H^{\omega}$$

Theorem 3. *Let $\sigma[f]$ be given by (2), and $b_k \geq 0$ ($k=1,2,3\ldots$)*

$$If\ \sum_{k=n}^{\infty}(B_k-\frac{b_k}{2k})=0(n^{-1})$$

then f is continuous of bounded variation.

Therem 4. *Let ω be a modulus of continuty, if $f\in H^{\omega}$, then $\psi\in H^{\omega}$.*

By this theorem H^{ω} class is invariant to ψ -transformation of Loo.

I.Javakhishili Tbilisi State University

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MATHEMATICS

G.Oniani

On Possible Meanings of Upper and Lower Derivatives

Presented by Academician L.Zhizhiashvili, November 27, 1995

ABSTRACT. **In the paper there is investigated the behaviour of upper and lower derivatives of the integrals with respect to the differentiation bases formed of intervals. The main result consists in that the analogy of the famous theorem of Besicovich generally is not true for translation invariant Busemann-Feller differentiation bases formed of intervals.**

1. A mapping B defined on $\mathbb{R}^n$ is called a differentiation basis in $\mathbb{R}^n$, if for every $x\in\mathbb{R}^n$ $B(x)$ is a collection of bounded open subsets of $\mathbb{R}^n$ containing x such that there is a sequence $\{R_k\}\subset B(x)$ with diam $R_k\to 0$ $(k\to\infty)$.

For $f\in L_{loc}(\mathbb{R}^n)$ numbers

$$\overline{D}_B\left(\int f,x\right)=\overline{\lim_{diam R\to 0, R\in B(x)}}\frac{1}{|R|}\int_R f \text{ and } \underline{D}_B\left(\int f,x\right)=\underline{\lim_{diam R\to 0, R\in B(x)}}\frac{1}{|R|}\int_R f$$

are called respectively upper and lower derivatives of the integral of f at a point x. If the upper and lower derivatives are equal, then their common meaning is called a derivative of the integral f, at a point x and it is denoted by $D_B\left(\int f,x\right)$. The basis B is said to differentiate the integral of f, if for almost every x

$$D_B\left(\int f,x\right)=f(x).$$

B is called a subbasis of B' (entry: $B\subset B'$) if $B(x)\subset B'(x)$ $(x\in\mathbb{R}^n)$. B is said to be a Buseman-Feller basis (BF - basis) if for every $R\in\overline{B}=\bigcup_{x\in\mathbb{R}^n}B(x)$ we have, that $R\in B(y)$for every $y\in R$. BF - basis is called homothety invariant (briefly: a HI - basis), if for $R\in\overline{B}$ $\overline{B}$ contains every set homothetical to R. Basis B is called translation invariant (briefly: a TI-basis), if $B(x)=\{x+R : R\in B(O)\}$ $(x\in R^n)$, where O is the origin of the coordinate system in $\mathbb{R}^n$.

Let B_2 denote the basis in $\mathbb{R}^2$, for which $B_2(x)$ $(x\in\mathbb{R}^2)$ consists of all twodimentional intervals containing x.

Let us agree $I^n=(0,1)^n$ and $f\in L(I^n)$ if $f\in L(\mathbb{R}^2)$ and $supp f\subset I^n$.

2. According to the famous theorem of Besicovitch (cf. [1] or [2], p.96) about upper and lower derivaties for every function $f\in L(\mathbb{R}^2)$ both of sets

$$\{x\in\mathbb{R}^2 : f(x)<\overline{D}_{B_2}\left(\int f,x\right)<\infty\},$$
$$\{x\in\mathbb{R}^2 : -\infty<\underline{D}_{B_2}\left(\int f,x\right)<f(x)\}$$

have a measure equal to zero.

Guzman and Menarges [2, p.100] established some generalization of this result. From their theorem follows that the analogy of Besicovich's statement is true for homothety invariant BF - bases $B\subset B_2$.

The question arises: may or not Besikovich's theorem be extended on translation invariant BF-bases $B\subset B_2$?

The negative answer on this question gives the following:

Theorem 1. *There is a translation invariant BF-basis $B\subset B_2$, for which there is a function $f\in L(I^2)$, $f\geq 0$ such that*

$$f(x)<\overline{D}_B\left(\int f, x\right)<\infty$$

almost everywere on I^2.

Here it must be noted: it is not difficult to prove that if $B\subset B_2$ is a translation invariant basis, then for every $f\in L(\mathbb{R}^2)$ inequalities are fulfilled almost everywere

$$\underline{D}_B\left(\int f, x\right)\leq f(x)\leq \overline{D}_B\left(\int f, x\right).$$

The question arises: if the analogous fact is true for all bases $B\subset B_2$?

The negative answer on this question gives:

Theorem 2. *There is a BF-basis $B\subset B_2$ for which there is a function $f\in L(I^2)$, $f\geq 0$ such that*

$$f(x)<\underline{D}_B\left(\int f, x\right)=\infty$$

almost everywhere on I^2.

Tbilisi State University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №1, 1996

MATHEMATICS

S.Topuria, N.Chikobava

Summability of the Fourier Series over Generalized Spherical Functions by the Abel and (C, α) Methods

Presented by Academician L.Zhizhiashvili, May 2, 1996.

ABSTRACT. **Some theorems on Abel and (C, α) summability of Fourier series over generalized spherical functions are proved.**

1. Notations and definitions. S^3 is a unit sphere on R^3 of the centre in origin of coordinates; (x, y) is scalar product of the vectors x and y; $D(x; h) = \{y: y \in S^3, (x, y) \geq \cos h, 0 < h \leq \pi\}$, $|D(x; h)|$ - area of the surface $D(x; h)$.

Definition 1. Let $f \in L(S^3)$. A point $x \in S^3$ will be called a L - point of the function f, if

$$\lim_{h \to 0} \frac{1}{h^2} \int_{D(x;h)} |f(y) - f(x)| ds(y) = 0.$$

Definition 2. A point $x \in S^3$ will be called a D - point of the function f, if

$$\lim_{h \to 0} \frac{1}{h^2} \int_{D(x;h)} [f(y) - f(x)] ds(y) = 0.$$

Definition 3. A point $x \in S^3$ will be called a $\widetilde{D}$ - point of the function f, if

$$\lim_{h \to 0} \frac{1}{h^2} \int_0^h \left| \int_{(x,y)=\cos\gamma} [f(y) - f(x)] dt(y) \right| d\gamma = 0.$$

Definition 4. A point $x \in S^3$ will be called a $L^*[D^*, \widetilde{D}^*]$ - point of the function f, if x and x^* are simultaneously $L[D, \widetilde{D}]$ - points of f; x^* being the opposite point to x.

It is known that if $f \in L(S^3)$ then almost every point of S^3 is its L^* - point.

The Fourier series over generalized spherical functions of the function $f \in L(S^3)$ is said to be the series

$$f(x) = f(\vartheta, \varphi) \sim \sum_{=|m|}^{\infty} I_\nu^{(m)}(f; \vartheta, \varphi), \tag{1}$$

Where

$$I_\nu^{(m)}(f; \vartheta, \varphi) = \frac{(-1)^m}{4\pi}(2\nu + 1) \int_{S^3} f(y) e^{-im(\varphi_1+\varphi_2)} \overset{\nu}{P}_{m,m}(\cos\gamma) ds(y) =$$

$$= \frac{(-1)^m}{4\pi}(2\nu + 1) \int_0^\pi \int_0^{2\pi} f(\vartheta', \varphi') e^{-im(\varphi_1+\varphi_2)} \overset{\nu}{P}_{m,m}(\cos\gamma) \sin\vartheta' d\vartheta' d\varphi',$$

$$\cos\gamma = \cos\vartheta\cos\vartheta' - \sin\vartheta\sin\vartheta'\cos\beta, \ \beta = \pi + \varphi' - \varphi, \ m = 0, \pm 1;$$

$$\mathrm{tg}\varphi_1 = \frac{\sin\beta\sin\vartheta'}{\cos\vartheta\sin\vartheta'\cos\beta + \cos\vartheta'\sin\vartheta}; \ \mathrm{tg}\varphi_2 = \frac{\sin\beta\sin\vartheta'}{\cos\vartheta'\sin\vartheta\cos\beta + \cos\vartheta\sin\vartheta'};$$

$$P^{\nu}_{m,n}(\mu) = \frac{(-1)^{\nu-m} i^{m-n}}{2^{\nu}(\nu-m)!}\sqrt{\frac{(\nu+m)!(\nu-m)!}{(\nu+n)!(\nu-n)!}}(1-\mu)^{\frac{n-m}{2}}(1+\mu)^{-\frac{m+n}{2}}\times$$

$$\times\frac{d^{\nu-m}}{d\mu^{\nu-m}}[(1-\mu)^{\nu-n}(1+\mu)^{\nu+n}],\ \mu=\cos\vartheta.$$

The system of generalized spherical functions coinsides with the system of common spherical functions [2] when $m = 0$, and (1) - with the Fourier - Laplaces series on sphere.

From the relation [3]

$$P^{\nu}_{1,1}(\mu) = P^{\nu}_{-1,-1}(\mu) = \frac{1}{2\nu+1}\cdot\frac{d}{d\mu}[p_{\nu+1}(\mu) - p_{\nu-1}(\mu)] - \frac{1}{2\nu+1}\cdot\frac{1}{1+\mu}[p_{\nu+1}(\mu) - p_{\nu-1}(\mu)],$$

due to [4].

$$\frac{d}{d\mu}[p_{\nu+1}(\mu) - p_{\nu-1}(\mu)] = (2\nu+1)p_{\nu}(\mu),$$

We obtain

$$P^{\nu}_{1,1}(\mu) = p_{\nu}(\mu) - \frac{1}{1+\mu}\int_{-1}^{\mu} p_{\nu}(t)dt,$$

i.e.

$$P^{\nu}_{1,1}(\cos\gamma) = p_{\nu}(\cos\gamma) - \frac{1}{1+\cos\gamma}\int_{\gamma}^{\pi} p_{\nu}(\cos t)\sin t dt, \tag{2}$$

where $p_{\nu}(\mu)$ is Legandres polynomial.

It is obvious from (2) that $P^{0}_{11}(\cos\gamma) = 0$. Hence, the series (1), when $m = \pm 1$, may be written as

$$f(\vartheta,\varphi) \sim \sum_{\nu=0}^{\infty} I^{(m)}_{\nu}(f;\vartheta,\varphi), \tag{3}$$

where

$$I^{(m)}_{\nu}(f;\vartheta,\varphi) = -\frac{2\nu+1}{4\pi}\int_0^{\pi}\int_0^{2\pi} f(\vartheta',\varphi')e^{-im(\varphi_1+\varphi_2)}[p_{\nu}(\cos\gamma) -$$

$$- \frac{1}{1+\cos\gamma}\int_{\gamma}^{\pi} p_{\nu}(\cos t)\sin t dt]\sin\vartheta' d\vartheta' d\varphi'.$$

2. **Summability of the Fourier series over generalized spherical functions by the Abel method**. The Abel means of (3) are denoted by

$$U(f;x) = U(f;\vartheta,\varphi) = \sum_{\nu=0}^{\infty} I^{(m)}_{\nu}(f;x)r^{\nu} =$$

$$= -\frac{1}{4\pi}\sum_{\nu=0}^{\infty}(2\nu+1)r^{\nu}\int_{S^3} P^{\nu}_{m,m}(\cos\gamma)f(y)e^{-im(\varphi_1+\varphi_2)}dS(y) =$$

$$= -\frac{1}{4\pi}\int_{S^3}\left[\sum_{\nu=0}^{\infty}(2\nu+1)p_{m,m}^{\nu}(\cos\gamma)r^{\nu}\right]f(y)e^{-im(\varphi_1+\varphi_2)}dS(y) =$$

$$= -\frac{1}{4\pi}\int_{S^3}p^*(r,\gamma)f(y)e^{-im(\varphi_1+\varphi_2)}dS(y),$$

where

$$p^*(r,\gamma) = \sum_{\nu=0}^{\infty}(2\nu+1)\,p_{m,m}^{\nu}(\cos\gamma)r^{\nu} =$$

$$= \sum_{\nu=0}^{\infty}(2\nu+1)p_{\nu}(\cos\gamma)r^{\nu} - \frac{1}{1+\cos\gamma}\int_{\gamma}^{\pi}\left[\sum_{\nu=0}^{\infty}(2\nu+1)p_{\nu}(\cos t)r^{\nu}\right]\sin t dt =$$

$$= p(r,\gamma) - p_1(r,\gamma).$$

$p(r,\gamma)$ is Poissons kernel to Fourier-Laplace series, i.e.,

$$p(r,\gamma) = \sum_{\nu=0}^{\infty}(2\nu+1)p_{\nu}(\cos\gamma)r^{\nu} = \frac{1-r^2}{(1-2r\cos\gamma+r^2)^{3/2}},$$

and

$$P_1(r,\gamma) = \frac{1}{2\cos^2\frac{\gamma}{2}}\int_{\gamma}^{\pi}P(r,t)\sin t dt.$$

We say that point $x(r,\vartheta,\varphi) \to x_0(1,\vartheta_0,\varphi_0)$ along paths nontangential to the sphere and write this circumstance as $x \xrightarrow{\Lambda} x_0$ if the point x tends to x_0 remaining all the time inside some cone with vertex at x_0 and angle $2\alpha < \pi$ whose axis coinsides with Ox_0 (O is a centre of a sphere).

The series (3) is called summable by the Abels method at the point $x_0(1,\zeta_0,\varphi_0)$ to the number S (or A - summable to the number S), if

$$\lim_{r\to 1-}U(f;r,\vartheta_0,\varphi_0) = S.$$

The series (3) is called summable by the A^* at the point x_0 to the number S if

$$\lim_{x \xrightarrow{\Lambda} x_0}U\ (f;r,\vartheta,\varphi) = S.$$

Theorem 1. *If $f(x)$ is continuous at some point $x_0(1,\vartheta_0,\varphi_0)$, then*

$$U(f;r,\vartheta,\varphi) \to f(\vartheta_0,\varphi_0)$$

howerer the point $x(r,\vartheta,\varphi)$ tends to x_0, remaining inside the sphere S^3.

Theorem 2. *If $f\in L(S^3)$, then series (3) is summable by the A^* method to $f(x)$ at all L - points of f.*

Theorem 3. *If $f\in L(S^3)$, then series (3) is summable by the Abel method to $f(x)$ at all D - points of f.*

3. **Summability of the Fourier series over generalized spherical functions by the (C,α), $\alpha>-1$, method**. Denote by $\sigma_n^{\alpha}(f;x)$ the Cesaro means (C,α), $\alpha>-1$, of series (3), i.e.

$$\sigma_n^\alpha(f;x)=\frac{1}{A_n^\alpha}\sum_{j=0}^{n}A_{n-j}^{\alpha-1}S_j(f;x)=$$

$$=-\frac{1}{2\pi}\int_0^\pi\int_0^{2\pi}f(\vartheta',\varphi')e^{-im(\varphi_1+\varphi_2)}\Phi_n^\alpha(\gamma)\sin\vartheta'd\vartheta'd\varphi',$$

where

$$\Phi_n^\alpha(\gamma)=K_n^\alpha(\gamma)-N_n^\alpha(\gamma),$$

$$K_n^\alpha(\gamma)=\frac{1}{A_n^\alpha}\sum_{j=0}^{n}\left(j+\frac{1}{2}\right)A_{n-j}^\alpha P_j(\cos\gamma),$$

$$N_n^\alpha(\gamma)=\frac{1}{2\cos^2\frac{\gamma}{2}}\int_\gamma^\pi K_n^\alpha(t)\sin t dt,$$

and $S_j(f;x)=S_j(f;\vartheta,\varphi)$ is the partial sum of series (3). $K_n^\alpha(\gamma)$ is a kernel of (C,α) means to Fourier - Laplace series [1].

Theorem 4. *Let $f\in L(S^3)$. Then the equality*

$$\lim_{n\to\infty}\sigma_n^\alpha(f;x)=f(x)$$

is fulfilled: 1) at all $\tilde{D}^*$ - points of f, if $\frac{1}{2}<\alpha<1$ and

2) at all $\tilde{D}$ - points of f, if $\alpha\geq 1$.

Georgian Technical University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №1, 1996

CYBERNETICS

Z.Piranashvili

On the Extension of Kotelnikoff-Shannon Formula

Presented by Academician V.Chavchanidze, June 10,1996.

ABSTRACT. **The extension of Kotelnikoff-Shannon formula which gives the possibility to increase interpolation step on the discreet points together with process value up to some order considering derivative value is given in the paper.**

The extension of results of the work [1] is given in the paper, namely it holds:

Theorem. If the conditions of theorem 1 in [1] is fulfilled, then for almost every sample distribution functions of random process $\xi(t)$ one has:

$$\xi(t)=\sum_{k=-\infty}^{\infty}\left\{\sum_{\tau=0}^{N}\frac{1}{(N-\tau)!\alpha^{N-r}}\left[\sum_{m=0}^{\tau}\frac{\left[\alpha\left(t-\frac{k\pi}{\alpha}\right)\right]^{N-m}}{(\tau-m)!}A_{m\tau n}\times\right.\right.$$

$$\left.\left.\times\sum_{j=0}^{m}\frac{\left(t-\frac{k\pi}{\alpha}\right)^{j}\xi^{(j)}\left(\frac{k\pi}{\alpha}\right)}{j!}\right]\varphi_{\tau N}(t;k,q,\alpha,\beta)\right\}\left[\frac{\sin\alpha\left(t-\frac{k\pi}{\alpha}\right)}{\alpha\left(t-\frac{k\pi}{\alpha}\right)}\right]^{N+1} \qquad (1)$$

for every $\alpha > \sigma/(N+1)$ and positive $\beta < [(N+1)\,\alpha - \sigma]/q$, where N and q are some fixed non-negative integers, and

$$A_{m\tau N}=\lim_{x\to 0}\frac{d^{\tau-m}}{dx^{\tau-m}}\left(\frac{x}{\sin x}\right)^{N+1},$$

$$\varphi_{\tau N}(t;\,k,\,q,\,\alpha,\,\beta)=\lim_{\varsigma\to\frac{k\pi}{\alpha}}\frac{d^{N-\tau}}{d\varsigma^{N-\tau}}\left(\frac{\sin\beta(\varsigma-t)}{\beta(\varsigma-t)}\right)^{q}.$$

The proof is similar to the one of theorem 1 in [1] and it is based on the following estimate: if $f(z)$ is integral function of exponential type with exponent σ, bounded on the real line, then for it the following estimater holds

$$\left|f(z)-\sum_{k=-n}^{n}\left\{\sum_{\tau=0}^{N}\frac{1}{(N-\tau)!\alpha^{N-\tau}}\left[\sum_{m=0}^{\tau}\frac{\left[\alpha\left(z-\frac{k\pi}{\alpha}\right)\right]^{N-m}}{(\tau-m)!}A_{m\tau N}\times\right.\right.\right.$$

$$\times\left.\left.\left(\sum_{j=0}^{m}\frac{\left(z-\frac{k\pi}{\alpha}\right)^{j} f^{(j)}\left(\frac{k\pi}{\alpha}\right)}{j!}\right)\right]\varphi_{\tau N}(z;k,q,\alpha,\beta)\right\}\left[\frac{\sin\alpha\left(z-\frac{k\pi}{\alpha}\right)}{\alpha\left(z-\frac{k\pi}{\alpha}\right)}\right]^{N+1}\Bigg| \leq$$

$$\leq\frac{L_f\cdot C_{qN}(z)}{[(N+1)\alpha-\sigma-q\beta]\cdot n^{q+1}}. \tag{2}$$

for every non-negative integers N and q and for any fixed z under sufficiently big n, where the function

$$C_{qN}(z)=\left(\frac{2\alpha}{\pi\beta}\right)^{q+1}\left(\frac{2}{1-e^{-\pi}}\right)^{N+1}\beta\cdot e^{q\beta|y|}\,|\sin^{N+1}\alpha z|$$

is finite in every bounded variation

$$z=x+iy,\ L_f=\sup_{-\infty<x<\infty}|f(x)|.$$

We obtain estimate (2) if for integral

$$\frac{1}{2\pi i}\int_{C_n}\frac{f(\zeta)}{\sin^{N+1}\alpha\zeta}\left(\frac{\sin\beta(\zeta-z)}{\beta(\zeta-z)}\right)^{q}\frac{d\zeta}{\zeta-z},$$

where C_n is circle $|\zeta|=(n+1/2)\,\pi/\alpha$, N and q are some non-negative integers, employ Cauchy theorem about residues and estimate the mentioned integral.

When $N=0$, then $\tau=0$, $m=0$ and from (1) we obtain formula [1]

$$\xi(t)=\sum_{k=-\infty}^{\infty}\xi\left(\frac{k\pi}{\alpha}\right)\frac{\sin\alpha\left(t-\frac{k\pi}{\alpha}\right)}{\alpha\left(t-\frac{k\pi}{\alpha}\right)}\left[\frac{\sin\beta\left(t-\frac{k\pi}{\alpha}\right)}{\beta\left(t-\frac{k\pi}{\alpha}\right)}\right]^{q}, \tag{3}$$

which is valid for every $\alpha>\sigma$ and positive $\beta<(\alpha-\sigma)/q$.

When $q=0$ from (3) we obtain known Kotelnikoff -Shannon formula

$$\xi(t)=\sum_{k=-\infty}^{\infty}\xi\left(\frac{k\pi}{\alpha}\right)\frac{\sin\alpha\left(t-\frac{k\pi}{\alpha}\right)}{\alpha\left(t-\frac{k\pi}{\alpha}\right)}, \tag{4}$$

where α is arbitrary fixed number more then σ.

When $N=1$, then from (1) we obtain formula

$$\xi(t)=\sum_{k=-\infty}^{\infty}\left\{\left[\left[(1+q)\frac{\sin\beta\left(t-\frac{k\pi}{\alpha}\right)}{\beta\left(t-\frac{k\pi}{\alpha}\right)}-q\cos\beta\left(t-\frac{k\pi}{\alpha}\right)\right]\xi\left(\frac{k\pi}{\alpha}\right)+\right.\right.$$

$$+\left(t-\frac{k\pi}{\alpha}\right)\frac{\sin\beta\left(t-\frac{k\pi}{\alpha}\right)}{\beta\left(t-\frac{k\pi}{\alpha}\right)}\xi'\left(\frac{k\pi}{\alpha}\right)\Bigg]\left[\frac{\sin\alpha\left(t-\frac{k\pi}{\alpha}\right)}{\alpha\left(t-\frac{k\pi}{\alpha}\right)}\right]^{2}\cdot\left[\frac{\sin\beta\left(t-\frac{k\pi}{\alpha}\right)}{\beta\left(t-\frac{k\pi}{\alpha}\right)}\right]^{q-1}, \qquad (5)$$

where $\alpha > \sigma/2,\ 0 < \beta < (2\alpha - \sigma)/q$.

When $q = 0$, then from (5) we obtain

$$\xi(t)=\sum_{k=-\infty}^{\infty}\left[\xi\left(\frac{k\pi}{\alpha}\right)+\left(t-\frac{k\pi}{\alpha}\right)\xi'\left(\frac{k\pi}{\alpha}\right)\right]\left[\frac{\sin\alpha\left(t-\frac{k\pi}{\alpha}\right)}{\alpha\left(t-\frac{k\pi}{\alpha}\right)}\right]^{2}, \qquad (6)$$

when $\alpha > \sigma/2$.

When $q = 0$, then from (1) we have

$$\xi(t)=\sum_{k=-\infty}^{\infty}\left\{\sum_{m=0}^{N}A_{mNN}\frac{\left[\alpha\left(t-\frac{k\pi}{\alpha}\right)\right]^{N-m}}{(N-m)!}\left(\sum_{j=0}^{m}\frac{\left(t-\frac{k\pi}{\alpha}\right)^{j}\xi^{(j)}\left(\frac{k\pi}{\alpha}\right)}{j!}\right)\right\}\times$$

$$\times\left[\frac{\sin\alpha\left(t-\frac{k\pi}{\alpha}\right)}{\alpha\left(t-\frac{k\pi}{\alpha}\right)}\right]^{N+1}$$

where $A_{mNN} = \lim\limits_{x\to 0}\frac{d^{N-m}}{dx^{N-m}}\left(\frac{x}{\sin x}\right)^{N+1}$. $\alpha > \sigma/(N+1)$ is any fixed number, N is some non-negative integer (when $N=0$, Kotelnikoff-Shannon formula (4) is obtained from (7)).

The formula (7) is given in [2] continuous functions with finite Fourier transform.

Observe that comparing formulae (1) and (7) with Kotelnikoff-Shannon formula, (1) and (7) gives possibility to increase the distance $h = \pi/2$ between points of interpolation $(N + 1)$-times and in particular cases for formulae (5) and (6) - twice. Moreover increasing q the speed of convergence of the above given expansions increased. Observe also that these expansions are valid for particular cases of spectral representation of covariance functions, namely for Karhunen-Loeve and Khinchin-Bochner representation. (Corollary 1,2,3 of theorem 1 [1]).

Institute of Cybernetics
Georgian Academy of Sciences

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PHYSICS

N.Kekelidze, R.Charmakadze, R.Alania, D.Kakushadze, A.Berdzenishvili, Z.Tkeshelashvili

Investigation of the Electroluminescent Properties of Gallium Nitride Structures Implanted with Zn and Mg Ions

Presented by Corr. Member of the Academy T.Sanadze, August 6, 1996

ABSTRACT. **Epitaxial GaN layers which always contain a great number of donor type defects were strongly compensated by doping in the growth process with further Zn and Mg ion implantation.**

Electroluminescent structures with new and important properties were obtained. Intensive light emission was produced in practically entire range of visible and near-ultraviolet light spectrum. The emission colour changes with variation in the degree of doping and the voltage applied to the structure. "Oscillation" of the current-voltage characteristic not observed in other structures was identified. Strong polarization of the short-wave band with a maximum at 3.29eV was found.

INTRODUCTION

Gallium nitride, a III-V semiconductor with a wide band gap (Eg=3.425 eV) [1] is very important and promising material. In particular, GaN light-emitting structures (LES) can be obtained for practically entire visible range including the near-UV region.

However today wide application of GaN is limited mainly due to the fact that GaN crystals and layers contain an extremely large number of native defects. Among these defects N vacances playing the role of donors are likely to be predominant [2]. The number of donor vacancies in GaN is so great that the electron concentration in undoped material may have the value $n=(1\div8)\cdot10^{19}cm^{-3}$. Owing to the indicated defective character of GaN, it is very difficult to obtain p-type material of adequate quality and thus a p-n junction which is so necessary for creation of highly efficient semiconductor devices. Therefore, at present, the development of production techniques of highquality GaN layers and crystals, their effective doping, etc. is of great importance. This is not feasible without fundamental investigations and the establishment of the character of defect formation in the material under discussion.

The present paper gives the results of technological and scientific investigations allowing construction of LES-s with unique characteristics on the basis of GaN.

EXPERIMENT

Radiative structures were grown on (1120) sapphire substrates using methods developed by Marushka [3] with some modifications in this regime.

Initially, GaN layers were grown on the substrate: they were undoped and thus had a very high carrier concentration of $8\cdot10^{19}cm^{-3}$. This layer is designated as n^{+}-GaN and is of 20μm thick (Fig. 1). Then during the growth in the chloridehydride flow Zn was introduced. This layer is designated as n-GaN. The layer thickness is 2÷3μm, $n=10^{18}cm^{-3}$. Then the Zn flow was interrupted and the substrate growth continued until

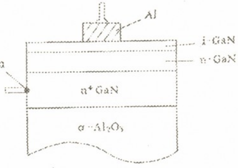

Fig.1. n^+-n-i structure of GaN on a sapphire substrate.

a layer of 0.8μm thick was obtained. The carrier concentration in n-GaN was approximately $3\cdot10^{18}cm^{-3}$.

An additional amount of Zn and Mg was introduced into the resultant structure by means of ion implantation. Profiles of zinc and magnesium distribution over the GaN layer thickness were recorded. The results are shown in Figs.2 and 3. All processes of the implantation of the gallium nitride layers were carried out in the same manner as in [4,5].

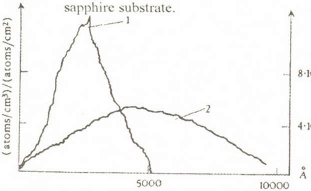

Fig.2. Depth distribution of Zn atoms in the epitaxial GaN layers (1 – before annealing, 2 – after annealing).

After Zn or Mg implantation, the GaN structures were subjected to multi-step annealing. Different techniques of annealing were used, such as photon, laser and temperature treatment. The typical disribution of Zn and Mg atoms over the thickness of the GaN epitaxial layers after annealing is shown in Figs.2 and 3 (curves 2).

Studies of zinc (magnesium) distribution conducted additionally by means of a scanning microscope showed the presence of a concentration gradient in the layers.

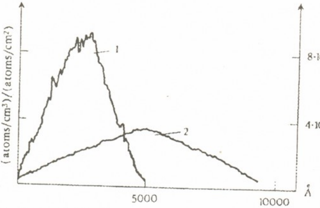

Fig.3. Depth distribution of Mg atoms in the epitaxial GaN layers (1 – before annealing, 2 – after annealing).

The quality of the layers was assessed by means of a diffractometer with GUR-5 goniometer and GP-4 attachment to DRON-1. The measurements showed a coarse-mosaic character of the epitaxial layers.

The mosaic character was studied by finding the difference between the reflections from a plane of (hh2he) and (hoho) type. The coarse mosaic character of the layers was identified by recording a doublet on the diffractometer at the frequency 1000 pulses/s with a counter rotational speed $\theta/2\theta \approx (1/16)$deg. at scanning (here θ is the Wolf-Bregg angle). The coarse-mosaic character of the layers was also revealed by X-ray using the Berg-Barret method in NiK_{β} line emission.

It should be noted that before ion implantation under the applied voltage of the polarities ("-" i-GaN) the structure emitted light in the wide spectral range, but for the given srtucture there is only one colour. The emission was unstable and of low intensity.

It should be emphasized that this structure can be essentially used as a source of near-UV emission with a maximum at 3.29 eV (300K). This happens with the change in the voltage polarity and when the concentration in the i-layer is of the order of $10^{17}cm^{-3}$. In such structures with near-UV emission, a high quantum efficiency of 0.3% is achieved.

RESULTS AND DISCUSSION

The end distribution of the charge carrier concentration (average) in the GaN layers before and after implantation and annealing is shown in Fig.4. The curves are obtained on the basis of measurements of electrical characteristics with the subsequent electrochemical etching of layers in combination with measurements of voltage-capacity characteristics. These measurements were made at certain structure depths after which the curves given in Fig.4 were constructed by means of approximation.Curve 4-1 corresponds to the initial impurity distribution, while curve 4-2 – to the distribution after implantation and annealing. The electron concentration

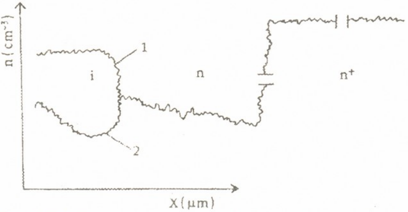

Fig.4. Free charge carrier concentration before implantation (1) and after implantation and annealing (2).

in n^+-GaN was $8{\cdot}10^{19}cm^{-3}$, the average electron concentration in the i-layer after annealing was $4{\cdot}10^{17}cm^{-3}$.

The peculiarity of the obtained structure is the presence of a compensated region (Fig.1 i-GaN) with a specific shape of the concentration gradient (curve 4-2).

When the voltage of 4V is applied (i.e. "+" on i-layer), the structure emits light from red to blue depending on the implantation degree and the i-layer thickness.

If the structure emits red light under the applied voltage, the increase of the voltage results in a sharp change first to yellow and then to green. If the structure emits yellow light, the increase of the voltage leads to green and then to blue light emission.

A typical emission spectrum for one of the samples is given in Fig.5. This blue emission band is characterized by quite a narrow spectrum half-width (of the order of 230 nm).

In such structures very high quantum efficiency (0.8-1.0%) is obtained for emission with a maximum at 2.89eV.

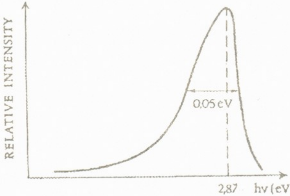

Fig.5. A typical emission spectrum for one of the samples.

A jumpwise character of electroluminescence is correlated somehow with the nature of I-V characteristics. Earlier, such voltage dependence of current was not observed in other structures. In particular, let us consider the structure with an "end" luminescence in the blue spectral region. Such structures, as a rule, begin to gleam at a voltage of the order of 4.5 V (point "a" in Fig.6). In this case the light is yellow (2.1 eV). With the further voltage increase, the

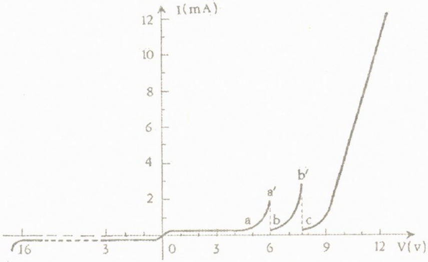

Fig.6. Current-voltage characteristic

yellow band intensity increases up to a certain level. When the voltage corresponds to point "a'", the luminescence actually goes out.

With a voltage increase from point "b", the structure begins to emit a green light (2.5eV), whose intensity rises up to the voltage corresponding to point "b'". At this point, luminescence actually goes out again.

A further voltage increase (from point "c") results in blue-light emission (2.87 eV); its intensity increases with a further increase in voltage. At large currents (of the order of 60 mA), a breakdown takes place in the structure.

To explain the processes occuring in electroluminescent structures various mechanisms, sometimes wholly or partially opposite to each other, are discussed in the literature [6-10]. Based on our experimental results and taking into account the date and considerations presented in [1-12], we think that when an electric field of sufficient strength is applied to the structure, electrons move from the n^+ layer to the n-layer and hence to the i-layer. On the other hand, under the influence of the electric field electrons flow from the i-layer into the metal. These electrons will originate mainly from Zn atoms in the proximity of the metal, whose energy levels lie in the forbidden region. In Fig.7 a schematic diagram (of the Zn levels) is shown: a) when no

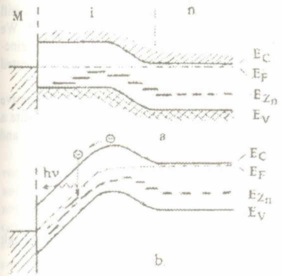

Fig.7. Schematic diagram

no voltage is applied, b) when voltage is applied. On the Zn level the remaining holes move into the i-layer depth.

It can be shown that the migration of holes in the i-layer requires a high electric field, that is why i-layer must be thin.

Electrons transferred from the n-layer into the i-layer recombine there with the holes resulting in the emission of visible light [8]. As it has been mentioned, ion implantation resulted in a concentration gradient, i.e. Zn levels are disposed at the various distance from the top of the Ev band.

When the external field of sufficient strength is applied, electrons passing from the n-layer into the i-layer recombine on the most deep Zn level. With increasing voltage the electron free path increases and hence the recombination already proceeds both energetically and geometrically on Zn levels situated in the different place. Exactly this circumstance causes the change in the emission spectrum towards the short-wave region with increasing voltage.

Production of descrete colours in a single structure was protected by the author certificate [10].

Taking the capacity-voltage characteristics of the emmiting crystal a diffusion length of carriers can be found. In our case it turned out to be approximately 1μm.

Initial GaN contains an abundance of native defects, mainly the vacances of nitrogen-V which are donors. We think that it is precisely these vacancies that play an important role both in undoped and doped gallium nitride as well as in degradation processes. In addition to nitrogen vacancies, there will be a very large number of bivacancies-2V in GaN.

When Zn or Mg atoms are introduced into GaN, they interact primarily with the vacancies forming the acceptor centre Zn+V. But the Zn+V centre will not be very stable.The activator complex Zn plus the bivacancy (Zn+2V) appears to be much more stable. Though Zn+V centres are formed in large numbers, they decay easier. Therefore, the concentration of Zn+2V complexes can be quite considerable, as well as their optical efficiency. Being introduced into a bivacancy, Zn forms a dumbbell-loke complex, i.e. an assymetric centre of dipole type.

Zn+2V+impurity (in particular Zn+2V+C, Zn+2V+0), Zn+V+impurity, Zn+3V should also be effective complexes. It is precisely these complexes that play a decisive role in the formation of intensive bands of GaN:Zn. It can be suggested here that the complex Zn+2V is resposible for the formation of the intensive band at 3.29eV characterised by a high degree of polarization. The complex Zn+2V+0 is probably associated with light emission at 2.55 eV.

Naturally, Ga vacancies, Zn and Ga interstitials, their substitutions and various associations with defects and impurities, which may appear in various processes, will be present in the material.

The polarization effect (60-70%) of the electroluminescent emission of GaN:Zn in the band with a maximum at 2.55 eV was first found in [13]. We have detected a still higher polarization (70-80%) in the short-wave band with a maximum at 3.29eV. We connect such high degrees of polarization with the presence of a great amount of zinc-bivacancy complexes of dipole type. Apparently, small regions of equally oriented dipole centres may also exist in the disordered GaN.

To conclude, it should be mentioned that the presence of great amount of native nitrogen vacancies as well as a high doping degree and implantation effects create a very complex situation manifesting itself in various properties of the material and requiring a further detailed investigation.

I.Javakhishvili Tbilisi State University

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PHYSICS

Z.Kachlishvili, K.Jandieri

Continuous Oscillations in Compensated Semiconductor under Impure Electric Breakdown

Presented by Corr. Member of the Academy N.Tzintzadze, June 1, 1996

ABSTRACT. **On the basis of mathematical model describing generation - recombination processes from impurity levels, dielectric relaxation of electric field and delay of electron temperature T_e relatively to the change of electric field necessary and sufficient condition of appearing persistant oscillations in compensated semiconductor under impure electric breakdown is established. This condition is reduced to rather simple form after a number of transformations. It contains the delay time of electron temperature τ_D as an important parameter characterizing the behavior of the system. The dependence of τ_D from T_e is studied. It is shown that while regarding the dynamics of the system τ_D with good accuracy can be considered constant.**

In the present paper we investigated the conditions of constant oscillations in extensively homogeneous, partially compensated semiconductor under impurity electric break down. The mathematical models used in [1,2,3] have been referred to. Relaxation equation average energy of charge carrier given in [1] was replaced by delay equation of electron temperature which allows to single out delay time τ_D as one of the most important parameters in determining system behaviour and to obtain simple analytic criterion of continuous oscillations beginning. Along with it examination of electron temperature delay unlike [2,3] gives the possibility to account the delay not only of the impact ionization coefficient A_I, but also the delay of μ mobility and thermal recombination coefficient B_T.

The dynamics of the system is characterized by the equations which describe:

1) generation and recombination process from impurity level;

$$\frac{dn}{dt} = (j\sigma + A_T)(N_D - N_A - n) + A_I(N_D - N_A - n)n - B_T(N_A + n)n \qquad (1)$$

2) dielectric relaxion of the electric field;

$$\frac{dE}{dt} = \frac{4\pi}{\varepsilon SR}(U - LE - e\mu SRnE) \qquad (2)$$

3) and delay of dimensionless electron temperature $Z = T_e / T$ relatively to the change of electric field (T is the temperature of lattice).

$$\frac{dZ}{dt} = \frac{Z - Z_0(E)}{\tau_D} \qquad (3)$$

Where N_D, N_A, n are donor concentrations, acceptors and free electrons correspondingly; $j\sigma$ and A_T are optic and thermal ionization rates from donor levels; E is electric field voltage; R is resistance of loading, switched in series with the sample; U is emf of continuous current source; L is the length of sample allong the field; S is the cross section of the sample; ε is dielectric penetration of the sample; z_0 is stationar meaning of z defined from the energy balance equation [4] for each fixed E, where along with phonon and ionized mechanism of energy relaxation unlike [1] the energy losses on $1S - 2P$ excitation of impurity atoms have been taken into consideration.

In order to study the system of equations (1) - (3) let's introduce new variables: $n = N_D X$, $U = E_B\zeta$, $E = E_B y$, $a_I = A_I N_D$, $b_T = B_T N_D$, where E_B is the meaning of the field in breaking down point.

According to Rauss - Hurvits criterion the equlibrium point of the system (x^*, y^*, z^*) is saddle - focus type point, if $pq - r < 0$, where p, q and r are the coefficients of corresponding characteristic equation $\Delta^3 + p\Delta^2 + q\Delta + r = 0$, which are always positive in our case. Then both continuous and chaotic oscillations can appear in the system. Thus, the condition of continuous oscillations is as follows:

$$\eta^2 + (2Wx^*\mu + F + \frac{1}{\tau_D} + \frac{z'_{0y}}{1+F\tau_D} Wx^*y^*\mu'_z)\eta +$$

$$+\frac{F}{\tau_D} + W\mu x^*\{\frac{1}{\tau_D} + \frac{z'_{0y}}{\tau_D(1+F\tau_D)} y^* \frac{\mu'_z}{\mu} +$$

$$+F + W\mu x^*(1 + \frac{z'_{0y}}{1+F\tau_D} y^* \frac{\mu'_z}{\mu}) - \frac{z'_{0y}}{1+F\tau_D} y^*(b'_z - a'_z x^*)\} < 0 \qquad (4)$$

where $\eta = 4\pi L/\varepsilon SR$, $W = 4\pi e N_D/\varepsilon$, $F = \sqrt{b^2 + 4ad}$, $b = -\gamma - b_T c + a_I(1-c)$, $a = a_I + b_T$, $d = \gamma(1-c)$, $\gamma = j\sigma + A_T$, $c = N_A/N_D$ - the compensation rate.

η and the coefficient standing before it are always positive. Therefore the fulfilment of (4) demands the negativity of free member. This final condition in its turn provides the fulfilment of (4) with the help of corresponding parameters L, S, R, ζ selection.

Thus the sufficient and necessary condition of the continuous oscillations is reduced to the negativity condition of free member (4). Calculations show that the last is carried only in rather narrow changing interval of E in the neighbourhood of break down point (as in [5,6]).

If scattering of impulse takes place on one of the defects (dynamic or static) then the corresponding mobility can be expressed in such a way: $\mu = \mu_0 z^m$ (where $m = -1/2$, 0, $3/2$ under scattering of impulse by acoustic phonons, by neutral and ionized atoms of impurity correspondingly). Let's regard the case of underlight absence; because of low tempratures we neglect thermal generator (i.e. $\gamma = 0$. It is easily seen that in this case the appearance of continuous oscillations is possible only when $E > E_B$ [3]). If we regard the behaviour of free member (4) in the neghbourhood of break down, the necessary and sufficient condition of continuous oscillations appearing will be as follows:

$$\tau_D b_T^B c(1 - \frac{1}{\beta}(y^* - 1)) > \frac{1+\alpha m}{2\alpha} \qquad (5)$$

where $\beta = \dfrac{\alpha}{1+\alpha m} \dfrac{a_I^B}{W\mu_0 z^{*m} c}$ and $\alpha = \dfrac{z'_{0y}(y^*)}{Z^*} y^*$ and index "B" indicates that the corresponding meaning is taken in break down point. Of course, to satisfy (5) it is necessary to fulfil the condition $y^* - 1 < \beta$, which defines this interval of changing y^*, in which continuous oscillations can appear. For the following analysis let's assume that $y^* - 1 = {}^{\beta}/_{2}$. Then if we take into consideration that according to our denotations $b_T^B = B_T^B N_D$, the condition (5) is reduced to:

$$\tau_D B_T^B N_D c > \frac{1 + \alpha m}{\alpha} \tag{6}$$

Thus , a rather simple criterion is observed, which allows to determine those meanings of parameters entered in it, for which the appearance of continuous oscillations is expected. As it is seen, among those parameters delay time τ_D plays an important role. The final part of the work is dedicated to the study of relation τ_D/τ_e (τ_e - is energy relaxation time) dependence on electron temperature. τ_e is defined from the equation: $KT(z_0 - 1)/\tau_e = g_{z0} = r_{z0}$, where $g = e\mu(z)E^2$ is acquisition power, and r is summed rate of energy loss [7]. The delay time τ_D is defined from the equation: $\tau_D = KT(z - z_0)/(r - g)$.

For different mechanisms of energy and impulse scattering we'll have the following various cases:

1. The energy is scattered on acoustic phonons: (for the corresponding rate of energy loss we use well-known Shokli's formula [8]), and impulse is scattered on phonones, on ionized and neutral atoms of impure. Correspondingly we'll have: (the notion $\Delta z = z - z_0$ is introduced)

$$\frac{\tau_D}{\tau_e} = \sqrt{1 + \frac{\Delta z}{z_0}} \left(2 - \frac{1}{z_0} + \frac{\Delta z}{z_0} \right)^{-1}; \quad \frac{\tau_D}{\tau_e} = Z_0 \left(1 + \frac{\Delta z}{z_0}\right)^{-1/2};$$

$$\frac{\tau_D}{\tau_e} = \left(1 + \sqrt{1 + \frac{\Delta z}{z_0}} \right) \left(2 + \sqrt{1 + \frac{\Delta z}{z_0}} + \frac{\Delta z}{z_0} - \frac{1}{z_0} \right)^{-1} \tag{7}$$

2. The energy is scattered on impurity atoms excitation [4]. For above given mechanisms of impulse scattering we'll have: (here and in the following case the results correspond to small deviations, when $\Delta z \le z_0$ are presented)

$$\frac{\tau_D}{\tau_e} \approx \frac{z_0 + \alpha}{(z_0 - 1)} \left(\frac{\alpha^2}{z_0^2} + \frac{\alpha}{z_0} + 1 \right)^{-1}; \quad \frac{\tau_D}{\tau_e} \approx \frac{z_0 + \alpha}{(z_0 - 1)} \left(\frac{\alpha^2}{z_0^2} - \frac{\alpha}{z_0} - 1 \right)^{-1};$$

$$\frac{\tau_D}{\tau_e} \approx \frac{z_0 + \alpha}{(z_0 - 1)} \left(\frac{2\alpha^2}{z_0^2} + \frac{\alpha}{z_0} + 1 \right)^{-1} \tag{8}$$

where $\alpha \equiv Tex/T$ (Tex is impure excitation temperature).

3. The energy is scattered on the ionization of impurity atoms

$$\frac{\tau_D}{\tau_e} \approx \frac{1}{(z_0 - 1)(A - B)}; \quad \frac{\tau_D}{\tau_e} \approx \frac{(A - B - 2/z_0)^{-1}}{(z_0 - 1)}; \quad \frac{\tau_D}{\tau_e} \approx \frac{(A - B - 1/2z_0)^{-1}}{(z_0 - 1)};$$

where $A \equiv 2\{\beta \ln[0{,}63(1 + 4z_0/\beta)]\,[1 + 4z_0/\beta + \sqrt{1 + 4z_0/\beta}\,]\}^{-1}$

$$B \equiv \beta E_i(-\beta/z_0)\{Z_0^2[\exp(-\beta/z_0) + \beta/z_0 E_i(-\beta/z_0)]\}^{-1}. \tag{9}$$

Here $\beta \equiv Tion/T$, where $Tion$ is impure ionization temperature.

For general case, when all enumerated mechanisms of impulse and energy scattering are regarded joinly, the dependence of relation τ_D/τ_e on electron temperature was obtained with the help of machine calculations in case of both small and arbitrary deviations. The obtained data show that approximate formulae describe the real dependence well if $\Delta z \leq 0,1z_0$. Under larger deviations for relation τ_D/τ_e they give high meanings. Figures 1 and 2 illustrate the dependence $\tau_D/\tau_e = f(z_0)$, obtained with the help of machine calculations in n - Ge at temperature $T = 4,2$ K for two different meanings of compensation rate and for different consentrations of donor impurity.

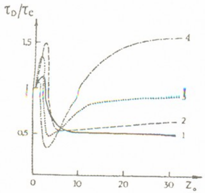

Fig. 1. τ_D/τ_e as a function z_0 in case of small deviations; $c = 0.9$, and N_D, takes the following meanings: 1.10^{13} cm^{-3}, 2.10^{14} cm^{-3}, 3.10^{15} cm^{-3}, 4.10^{16} cm^{-3}

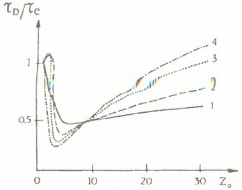

Fig. 2. τ_D/τ_e as a function z_0 in case of small deviations; $c = 0.1$, and N_D, takes the following meanings: 1.10^{13} cm^{-3}, 2.10^{14} cm^{-3}, 3.10^{15} cm^{-3}, 4.10^{16} cm^{-3}

At the end it should be noted: that oscillations appear in narrow interval of changing z^* in the neighbourhood of breakdown point z_B, $z_B >> 1$, if only the smallest meanings of compensation are not regarded. Besides the amplitude of oscillation is small relatively to z^*. Proceeding from this $\Delta z << z_0$. As it is seen from the given formulae when $\Delta z >> 1$ and $\Delta z << z_0$, the relation τ_D/τ_e is slightly depended on z and z_0 (the given Figures testify it also). Besides in the marked interval the change of τ_e is slight. All this gives the ground while regarding the dynamics of the system to consider τ_D as constant parameter (for the given C and N_D).

I.Javakhishvili Tbilisi State University

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PHYSICS

E.Khutsishvili, M.Paghava, A.Kuchashvili

Thermal Expansion Coefficient of Gallium Arsenide at High Temperatures

Presented by Corr. Member of the Academy G. Tsagareishvili, June 22, 1996

ABSTRACT. **Thermal exspansion coefficient of single and polycrystals gallium arsenide has been investigated at the temperature range from 300 to 900 K.**

Good agreement of gallium arsenide thermal expansion coefficient temperature dependence with that of heat capacity of semiconductors was found.

The effect of annealing at 823 and at 873 K of gallium arsenide is expected to be connected with the process of arsenic impoverishment of the material.

Gallium arsenide for its physical properties is widely used for many fields of semiconductor technical applications. One of the most necessary characteristics applied material in the fabrication of devices is the thermal expansion coefficient.

Few reports, however, have been published on the thermal expansion coefficient of gallium arsenide [1-7]. There are some discrepancies in the data of references. At the same time data given in references [1-7] belong to the low temperatures (< 300K). It will be necessary to extend the measurements of thermal expansion coefficient of gallium arsenide to higher temperatures.

In the present paper thermal expansion coefficient (α) of gallium arsenide single crystals (with [III] orientation) and polycrystals at the high temperatures in the range 300-900K have been investigated. We have carried out measurements in air using quartz dilatometer. The measurements were performerd on gallium arsenide samples before and after thermal annealing at 823K (for 2 hrs) and 873K (for 1 hr) in hydrogen atmosphere. The experimental temperature dependence of thermal expansion coefficient of gallium arsenide is plotted in Figure.

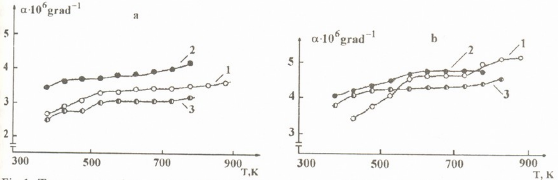

Fig.1. Temperature dependence of thermal expansion coefficient of gallium arsenide single crystals (a) and polycrystals (b); 1 - initial state; 2 - annealing at 823K two hrs; 3 - annealing at 873K one hr

The increase of thermal expansion coefficient of gallium arsenide with temperature is the same for both single crystals and polycrystals: in the range 300-600K α increases relatively faster than at higher temperatures, where this increase is much less and α approaches to the saturation. This dependence is similar to temperature change of heat capacity characteristic of semiconductors over the whole investigated temperature range.

As to the annealing our experimental results have shown that the influence of annealing on the thermal expansion coefficient of gallium arsenide polycrystals is not observed. However first annealing causes weak increase of thermal expansion coefficient of gallium arsenide single crystals. After the second annealing the results of thermal expansion coefficient approach to the values of unannealed samples.

If we assume that the obtained results are connected with gallium arsenide belonging to chemical compounds with break structure we can presume that the break annealing causes arsenic losses from the material.

F.Tavadze Institute of Metallurgy
Georgian Academy of Sciences

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
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PHYSICS

A.Rukhadze, M.Chogovadze, L.Chachkhiani

Radiative Instability in Ideal Liquid with the Tangential Discontinuity

Presented by Corr. Member of the Academy N. Tsintsadze, June 27, 1996

ABSTRACT. **The nature of the tangential discontinuity in the ideal liquid has been investigated. New type of instability - radiative instability stipulated by motion of a light liquid above the surface of heavy one has been discovered. This radiative instability should be considered as stimulated Cherenkov sound wave radiation from discontinuity surface.**

The theory of instability of the tangential discontinuity in the ideal liquid [1] is generalized by accounting the radiation from its surface. Contrary to the well-known instability of discontinuity related to the surface waves excitation by the inhomogeneity of stream velocity (Kelvin-Helmholts instability [2]) the considered radiative instability is stipulated by the supersonic motion of a light liquid above the surface of heavy one.

1. The equation of motion of the ideal liquid [1]

$$\operatorname{div} H = 0, \quad \operatorname{rot}[\mathbf{v}\mathrm{x}\mathbf{H}] = \frac{\partial \mathbf{H}}{\partial t}$$

$$\frac{\partial \mathbf{v}}{\partial t} + (\mathbf{v}\nabla)\mathbf{v} = -\frac{\nabla P}{\rho} + \frac{1}{4\pi\rho}[\operatorname{rot}\mathbf{H}\times\mathbf{H}] \qquad (1)$$

$$\frac{\partial \rho}{\partial t} + \operatorname{div}\rho\mathbf{v} = 0, \qquad \frac{\partial P}{\partial \rho} = c_o^2$$

admit the solution with inhomogeneous $\rho_o(x)$, $P_o(x)$, $\mathbf{v}_o(x)$ and $\mathbf{H}_o(x)$ when $\mathbf{v}_o \parallel \mathbf{H}_o \parallel OZ$ and the relation

$$P_o(x) + \frac{H_o^2(x)}{8\pi} = \text{const} \qquad (2)$$

is satisfied. However it is well-known that such motion in general occurs to be unstable. Usually two types of instability are considered: the volumetric one stipulated by the smooth inhomogeneities of the equilibrium parameters of liquid and surface one which takes place when inhomogeneities are sharp and there exists a discontinuity surface in the liquid. In the first case the system of equations (1) for the small volumetric perturbations is analyzed with the given boundary conditions, whereas in the second case the surface type perturbations located near the tangential discontinuity are considered. It will be shown below that these two types of instabilities have one nature and at the same time the second one represents the limiting case of the first.

Besides the theory of instability tangential discontinuity is generalized by accounting the radiation from the discontinuity surface. Such instability takes place when the light liquid is moving with supersonic velocity above the surface of heavy liquid. It is obvious that it can be considered as stimulated Cherenkov radiation of sound waves from the surface of discontinuity.

2. The most complete analysis of the instabilities of tangential discontinuity related to the excitation of surface waves was done in the paper by V.G.Kirtskhalia [3]. In accordance to this paper the system of equations (1) can be reduced to one equation for quantity

$$y = \frac{v_{1x}}{\omega - k_z v_o}$$

Supposing that $y \sim \exp(-i\omega t + ik_z z)$ this equation will be

$$\frac{d^2y}{dx^2} + \{\frac{[(\omega - k_z v_o)^2 - k_z^2 v_A^2][(\omega - k_z v_o)^2 - k_z^2 c_o^2]}{(\omega - k_z v_o)^2 (v_A^2 + c_o^2) - k_z^2 v_A^2 c_o^2}\} y = 0, \tag{3}$$

where v_o is the stream's velocity, c_o is the sound velocity, $v_A = H/(4\pi\rho)^{1/2}$ is Alfen speed.

This equation is valid in both sides of the discontinuity surface $x = 0$, on which the quantities c_o^2, v_A^2, and v_o are jumping, but the relation (2) takes place.

The boundary conditions for (3) can be obtained by integration of equations (1) near the surface $x = 0$, assuming that the quantities **H** and ρ are finite:

$$\{y\}_{x=0} = 0, \quad \{\frac{(\omega - k_z v_o)^2 (v_A^2 + c_o^2) - k_z^2 v_A^2 c_o^2}{(\omega - k_z v_o)^2 - k_z^2 c_o^2} y'\}_{x=0} = 0. \tag{4}$$

The notation $\{A\}_{x=o}$ means the jump of A at x=0.

Supposing that $y(x) \to 0$ at $x \to \pm\infty$ from (3) and (4) in the paper [3] the following dispersion equation was obtained

$$\beta_1 \chi_2 + \beta_2 \chi_1 = 0 \tag{5}$$

$$\chi_i^2 = \frac{[(\omega - k_z v_{oi})^2 - k_z^2 c_{oi}^2][(\omega - k_z v_{oi})^2 - k_z^2 v_{Ai}^2]}{k_z^2 c_{oi}^2 v_{Ai}^2 - (\omega - k_z v_{oi})^2 (v_{Ai}^2 + c_{oi}^2)},$$

$$\beta_i = (\omega - k_z v_{oi})^2 - k_z^2 v_{Ai}^2 \tag{6}$$

where $(i = 1,2)$

In [3] it was supposed that $\chi_i^2 > 0$ that corresponds to the exponentially damping solutions y(x) at $x = \pm\infty$. However, it is easy to show that the equation (5) stays valid even when χ_i^2 is complex with $\mathrm{Re}\chi_i^2 > 0$. The solutions y(x) then correspond to the volumetric waves propagating from the discontinuity surface at $x = 0$. Such solutions are considered below.

3. Let us remind some results of instability theory of tangential discontinuity for the surface waves, i.e. when $\chi_i^2 > 0$. We'll restrict by the simplest case when only the stream's velocity is gamping, or

$$v_o(x) = \begin{cases} +v_o, & x > 0, \\ -v_o, & x < 0. \end{cases} \tag{7}$$

For noncompressing liquid, when $c_o^2 \to \infty$ (more correspondingly $c_o^2 >> v_A^2, v_o^2$) from (6) it follows that $\chi_i^2 = k_z^2 > 0$. The dispersion equation (5) then has the exact solution

$$\omega^2 = k_z^2 (v_A^2 - v_o^2). \tag{8}$$

We see that this solution corresponds to the unstable oscillations ($\omega^2 < 0$) if $v_o^2 > v_A^2$. Thus, in the absence of external magnetic field, $\mathbf{H}_o \to 0$, the tangential discontinuity occurs to be unstable [2]. The external magnetic field with $v_A^2 > v_o^2$ stabilizes this instability, as it was first shown by S.I.Syrovatsky [4].

Strictly speaking all the above statements are correct only for the surface type oscillations. Now we also generalize the results for the volumetric oscillations.

4. For clarifying the nature of considered instability we analyze again the stability of smoothly inhomogeneous, noncompressing liquid system with velocity profile $v_o(x)$ in the absence of external magnetic field. Then from (1) in linear approximation one can easily obtain

$$\frac{d^2 v_{1x}}{dx^2} - \left[k_z^2 - \frac{k_z v_o''(x)}{\omega - k_z v_o(x)}\right] v_{1x} = 0. \tag{9}$$

This equation must be completed by the following boundary conditions

$$v_{1x}\Big|_{x=\pm a} = 0. \tag{10}$$

First of all it is obvious, that the instability is possible only if in the considered region $-a \le x \le a$ there exists a point where $v_o''(x) = 0$. But this condition is not satisfactory. For example, if the stream velocity $v_o(x)$ is symmetric, the instability is impossible. Really if $v_o(x) = u_o x^2/a_o^2$ (the homogeneous part of $v_o(x)$ can be easily excluded) then from (9) the longwave perturbations, $|k_z|a \ll 1$, it follows that

$$v_{1x}'' - \frac{2}{x^2 - \dfrac{\omega a_o^2}{k_z u_o}} v_{1x} = 0. \tag{11}$$

Introducing $\xi = x(k_z u_o / \omega a_o^2)^{1/2}$ and $u = v_{1x}'$ the equation (11) can be reduced to the standard Legandre equation

$$(\xi^2 - 1)u'' + 2\xi u' - 2u = 0 \tag{12}$$

the general solution of which looks like

$$u = C_1 P_1(x(k_z u_o/\omega a_o^2)^{1/2}) + C_2 Q_2(x(k_z u_o/\omega a_o^2)^{1/2}) \tag{13}$$

Substituting this solution into the boundary condition (10) we obtain the dispersion equation describing time evolution of small initial perturbations. It is easy to show that this dispersion equation has only real solutions $\omega(k_z)$. This means that the perturbations are stable.

Quite another situation arises if $v_o(x)$ is an old function of x with $v_o''(x) \neq 0$ in the region $-a \le x \le a$. Thus if $v_o(x) = u_o x^2/a_o^2 + \beta x$ then from (9) one can obtain the standard Bessel equation

$$v_{1x}'' - \left\{k_z^2 - \frac{2}{[x + (-\omega a_o^2/k_z u_o)^{1/2}]^2}\right\} v_{1x} = 0, \tag{14}$$

where $a_o^2\beta/2u_o \equiv (-\omega a_o^2 / k_z u_o)^{1/2}$. This equation was very carefully investigated in [5]. It was shown that it has unstable solutions with $\text{Im}\,\omega = k_z u_o[(3)^{1/2}/6]a^2/a_o^2$. Namely this instability corresponds to the well-known Kelvin-Helmholts instability [2] and represents the genesis of the surface waves instability considered above.

5. Finally let us consider the possibility of existence of the radiative instability in the ideal liquid with tangential discontiniuty. For simplicity we restrict ourselves by the consideration of the situation when a light liquid stream is moving above the surface of heavy one with velocity profile

$$v_o(x) = \begin{cases} v_o, & x > 0, \\ 0, & x < 0. \end{cases} \quad (15)$$

and the external magnetic field absents. Then in accordance with equilibrium $c_{01}^2 >> c_{02}^2$. Therefore the light liquid can be considered as noncompressible, $c_{01}^2 \to \infty$. As a result we obtain $\chi_1^2 = k_z^2 > 0$. This means that in the light liquid the perturbations occur to be of screening, or surface type. On the contrary in the heavy liquid they occur to be volumetric with $\mathrm{Re}\,\chi_2^2 < 0$ and slightly damping $0 < \mathrm{Im}\chi_2^2 << |\mathrm{Re}\chi_2^2|$. Then from the dispersion equation (5) one can easily find the following solutions

$$\omega = k_z v_o \pm i\,|k_z|\,c_{o2} \quad (16)$$

The solution with $\mathrm{Im}\omega > 0$ corresponds to the unstable perturbations. It is easy to understand that this instability represents the stimulated Cherenkov radiation of sound waves exciting by the light liquid stream, moving above the surface of heavy liquid.

In conclusion it should be noted, that Cherenkov type instability as the above considered (8) can be stabilized by the action of external magnetic field if $v_A^2 > c_o^2$.

Georgian Technical University

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ANALYTICAL CHEMISTRY

O.Manjgaladze, M.Gverdtsiteli

Application of Theory of Graphs in Systemic-Structural Analysis of Educational Material

Presented by Academician G.Tsintsadze, June 17, 1996

ABSTRACT. **The method of systemic analysis is considered. The mathematical criterion of efficiency of this method, relative information, is devised in terms of graphs theory.**

Modern pedagogics pay particular attention to systemic approach of educational process. The methods of systemic analysis,i.e. logical structuration and logical dosage were elaborated by O.Zaitsev [1,2] and E.Mishina [3]. According to this approach the concepts are used as structural elements and logical connections between them as system-generators for educational material. Three types of logical connections are distinguished: genetic, subordinate and co-subordinate.

The genetic connections indicate the essence of event and stages of its development. For example: oxidizer (1) promotes process of oxidation (2). The connection (1)-(2) is genetic and is directed from cause (1) to effect (2).

The subordinate connections are directed from family concept to species concept. For example: chemical reactions (3) are devided into combination (4), decomposition (5), substitution (6) and metathesis (7) processes. (3)-(4), (3)-(5), (3)-(6) and (3)-(7) connections are subordinate.

Co-subordinate connections exist among the concepts of one family and they have no directions. In our example co-subordinate connections are: (4)-(5), (4)-(6), (4)-(7), (5)-(6), (5)-(7) and (6)-(7).

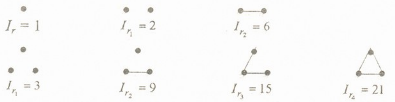
This method was modernized and applied to construct the course of theoretical principles of analytical chemistry by O.Mandjgaladze [4,5].

Mathematical criterion of efficiency of the logical-structural method, the concept of relative information (I_r), was elaborated in terms of theory of the graphs [6]:

$$I_r = N(\sum_i \deg V_i + 1)$$

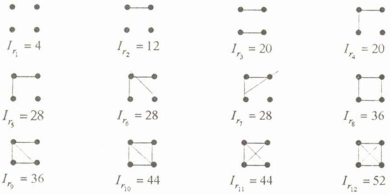
where: N is a number of vertexes in graph (the number of structural elements in the system); $degV_i$ - the degrees of vertexes of graph (numbers of logical connections among structural elements).

The values of I_r for one-, two-, three- and four-term systems are brought below:



$I_r = 1$ $I_{r_1} = 2$ $I_{r_2} = 6$

$I_{r_1} = 3$ $I_{r_2} = 9$ $I_{r_3} = 15$ $I_{r_4} = 21$



As we see for two-term system in a case of maximum connection (full-graph) I_{r_2} is 3 times as large as in case of non-connection (zero-graph); for three-term system I_{r_4} / I_{r_1} =7; for four-term system $I_{r_{12}} / I_{r_1}$ =13. So, I_r increases with increase of structural elements numbers in system and numbers of logical connections among them. This result is universal for every N.

By modernizing this approach I_r can be calculated for the systems with non-equivalent concepts (painted graphs) and non-equivalent connections (orientaited graphs).

I.Javakhishvili Tbilisi State University

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ORGANIC CHEMISTRY

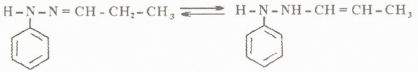
M.Gverdtsiteli, N.Samsonia

Algebraic Investigation of Fisher's Reaction

Presented by Corr. Member of the Academy L.M.Khananashvili, June 29, 1995

ABSTRACT. **The algebraic method of recording of organic molecules and reactions in terms of ANB - matrices is presented. Their diagonal elements are the atomic numbers of chemical elements, whereas nondiagonal ones - the multiplicities of chemical bonds. Algebraic investigation of Fisher's reaction is given in terms of this method.**

Fisher's reaction is an important method for synthesis of homologues of indole. The main stage of this complex process (according Robinson) is rearrangement of tautomeric form of phenylhydrazone, which proceeds analogous to benzidine rearrangement [1]:



Theoretical investigation of this rearangement was carried out in terms of algebraic chemistry [2].

Contiguity matrices and their various modifications are efficiently used in modern theoretical organic chemistry for characterization of molecules and their transformation [2,3]. One type of such matrices are ANB - matrices. Their diagonal elements represent atomic numbers of chemical elements, whereas nondiagonal ones - the multiplicities of chemical bonds [4]. For arbitrary ABC molecule ANB - matrix has a form:

$$\overset{1}{A}\overset{2}{B}\overset{3}{C} \qquad \left\| \begin{matrix} Z_A & \Delta_{AB} & \Delta_{AC} \\ \Delta_{AB} & Z_B & \Delta_{AC} \\ \Delta_{AC} & \Delta_{BC} & Z_C \end{matrix} \right\|$$

where: Z_A, Z_B, Z_C are atomic numbers of A, B and C chemical elements; Δ_{AB}, Δ_{AC}, Δ_{BC} represents multiplicities of chemical bonds between A and B, A and C, B and C.

The modelling reaction of phenylhydrazone rearrangement and its notation in ANB - matrices form are brought below:

$$H_2N{-}N = CH{-}CH3 \rightleftarrows H_2N{-}NH{-}CH = CH_2$$

$$\begin{Vmatrix} 7 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 1 \\ 1 & 7 & 2 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 2 & 6 & 1 & 0 & 0 & 0 & 1 & 0 & 0 \\ 0 & 0 & 1 & 6 & 1 & 1 & 1 & 0 & 0 & 0 \\ 0 & 0 & 0 & 1 & 1 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 1 & 0 & 1 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 1 & 0 & 0 & 1 & 0 & 0 & 0 \\ 0 & 0 & 1 & 0 & 0 & 0 & 0 & 1 & 0 & 0 \\ 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 0 \\ 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 \end{Vmatrix} \longrightarrow \begin{Vmatrix} 7 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 1 \\ 1 & 7 & 1 & 0 & 1 & 0 & 0 & 0 & 0 & 0 \\ 0 & 1 & 6 & 2 & 0 & 0 & 0 & 1 & 0 & 0 \\ 0 & 0 & 2 & 6 & 0 & 1 & 1 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 & 1 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 1 & 0 & 1 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 1 & 0 & 0 & 1 & 0 & 0 & 0 \\ 0 & 0 & 1 & 0 & 0 & 0 & 0 & 1 & 0 & 0 \\ 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 0 \\ 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 \end{Vmatrix}$$

Let's consider the expression;

$$\Delta_r = \Delta_f - \Delta_i$$

where: Δ_i - is the value of determinant of ANB - matrix for the initial form; Δ_f - for final form; Δ_r - change of the value of determinant in the result of rearangement.

As calculations show:

$$\Delta_r = 444 - 416 = 28$$

Thus, the algebraic criterion of this process is changing the determinant value of ANB - matrices.

I.Javakhishvili Tbilisi State University

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ORGANIC CHEMISTRY

M.Sturua, T.Vephkhvadze, R.Ziaev, D.Tsakadze, Sh.Samsonia, A.Abdusamatov

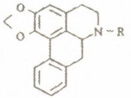
Alkaloids of Magnolia (*Magnolia obovata Thunb*)

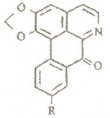
Presented by Corr. Member of the Academy D.Ugrekhelidze, July 21, 1996

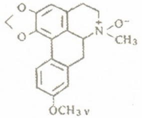
ABSTRACT. **From magnolia (*Magnolia obovata Thunb*) gathered in Batumi Georgia) well-known alkaloids remerin, anonain, liriodenin and lanuginozin were isolated and identified. By modern methods the structure of fifth alkaloid was ascertained - N-oxide of isolaurelin, which is isolated from plants for the first time.**

The plant from genus *Magnolia* is related to family of *Magnoliaceae* and is represented by about 70 species. It is known that genus Magnolia contains considerable amount of alkaloids. Bis-benzylisokhnolic [1], aporphic alkaloids and their oxo - derivatives [2,3] were isolated from some species.

Leaves of *Magnolia obovata Thunb* (gathered in Batumi) were investigated. By extraction with chloroform 0.48% sum of alkaloids were received. It was devided in two parts - phenolic and nonphenolic. From phenolic part lireodenin [1] was isolated. On chromatographic column with silicagel from mother liquor - anonain (II), remerin (III), liriodenin (I), lanuginozin (IV) and new base (V) were isolated.







II R = H
III R = CH_3

I R = H
IV R = OCH_3

Base (V) has composition $C_{19}H_{19}NO_4$. Melting point is 138 - 140^{0}C (acetone). This base is low - soluble in organic solvents and high - soluble in water. IR spectrum of (V) contains maximums of absorption at 223, 283 nm (lg ε 4.35; 4.18); in PMR - spectrum signals appear at: from protons of N - methylic group (2.90 m.p. 3H, singlet), methoxyl - (3.86 m.p., 3H singlet), methylendioxi - (5.95 and 6.10 m.p. single - proton dublets with J = 1.5 Hz), and signals of four aromatic protons. In mass-spectrum there are peaks with m/z 325 (M^+, 3.5%), 309 $(M - 16)^+$, 308 $(M - 17)^+$, 307 $(M - 18)^+$, 294 $(M - 31)^+$, 292, 267, 266 (100%), 265, 251, 235.

High solubility in water, characteristic shift of signals of N-methyl groups protons to faint field [4], faint intensity if peak of molecular ion and presence in mass - spectrum of (V)peaks of ions $(M - 16)^+$, $(M - 18)^+$ indicate that base (V) is N-oxide of isolaurelin [5]. Reduced by zinc in 10% of muriatic acid, compound identical to isolaurelin was isolated.

In this way five bases were isolated from leaves of *Magnolia obovata Thunb*. N-oxide of isolaurelin is found in plants for the first time.

In chromatography silicagel L (Czechoslovakia) 5/40 mkm was used with addition of 5% of gypsum (as a fixative) - TLC, 40/100 mkm (column chromatography). Benzen-methanol (4:1) was used as a solvent. IR spectrum were taken on spectrometer UR - 20 in KBr: PMR - on JNM - 4H - 100/100 mHz (δ - scale, CD_3OD, MDS); mass-spectrum - on instrument MX - 1310.

Isolation of alkaloids. 2 kg of dry leaves of *Magnolia obovata Thunb* were soltened by 5% solution of ammonia and alkaloids were extracted by chloroform. Extract was treated by 10% solution of sulphuric acid. The solution was alkalizated by 25% solution of ammonia and alkaloids were comprehensively extracted by chloroform. 9.6 g or 0,48% sum of alkaloids from dry mass was obtained.

The alkaloids sum was treated by 10% of sulphuric acid. The solution was alkalizated by 25% of ammonia and alkaloids were extracted by ether. Extract (0,5 l) was treated by 5% solution of caustic soda and washed by water. After destilling ether 5.85 g of nonphenolic part of alkaloids sum was isolated. Alkalizated solution was accidified by muriatic acid (1:1), again alkalizated by 25% solution of ammonia and extracted by ether. 2.1 g of phenolic bases were obtained.

Nonphenolic fraction was treated by chloroform and 0,9 g of liriodenin was isolated. Mother liqour was chromatographed on column with silicagel. By eluting with mixture benzene - methanol (99:1, 98:2, 95:5), anonain (0.056 g), remerin (0.15 g), liriodenin (0.22 g), lanuginozin (0.15 g) and N-oxide of osolaurelin (0.08 g) were isolated.

Well known alkaloids were identified by direct comparison with variable samples of alkaloids, isolated from *Liriodendron tulipifera* L from family *Magnoliaceae*.

N-oxide of isolaurelin. By treatment of benzene - methanol eluate (95:5) with aceton the base with composition $C_{19}H_{19}NO_4$ was isolated. Melting point 138 - 140^{0}C, Rf 0.18.

Reduction of N-oxide of isolaurelin. 0.04 g of base was isolated in 10% solution of sulphuric acid; zinc dust was added and kept for twenty four hours. The mixture was filtrated and alkalizated by 25% of amonia and extracted by ether. By elutriation of ether the base was isolated with Rf 0.64, equal to Rf of isolaurelin.

I.Javakhishvili Tbilisi State University Tashkent Agricultural Institute

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ORGANIC CHEMISTRY

M.Kezherashvili, N.Lekishvili, L.Asatiani, B.Butskhrikidze, L.Kipiani

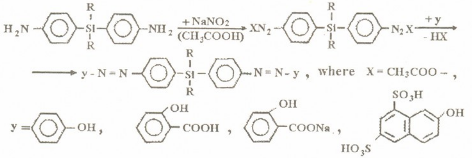
The Silicon-Containing New Bisazodyes

Presented by Corr. Member of the Academy L.Khananashvili, May 20, 1996

ABSTRACT. **The silicon-containing new bisazodyes were synthesized and studied. Some of the synthesized dyes were investigated for dyeing natural silk fibres. It was established that the colours of obtained dyes on natural silk is characterized by brightness and high fastness to different treatment.**

It is known from literature that introduction of dialkyl (aryl) silil groups into molecular composition of organic azodyes increases their thermostability and improves their hydrophobic properties, etc. [1]. We carried out synthesis of new bisazodyes containing silicium atoms, which have been based on the Bis - (Para - aminophenil) silane as initial diamine [2].

We obtained bisazodyes by the reaction of tetranitrogenation and by the subsequent azocoupling of the obtained tetraazonium salt with various azoic coupling components. The reaction may be expressed by the following scheme:



We conducted the reaction of tetranitrogenation in the area of acetic acid at about 0-5°C temperature, according to the method described in [3]. For example, we added previously prepared suspension (0.5 mole initial diamine in the 300 ml water) the 0.2 mole acetic acid as 25% (per) aqueous solution of 0.1 mole $NaNaO_2$. We continued to stir it for about 40-45 minutes, added it 0.1 mole sodium salt of salicile acid, 0.25 mole $NaHCO_3$ and stirred reaction mixture for about 60 minutes. At the end of reaction we filtered and washed the educt sediment by cold water. We purified the obtained product using extraction method in the Sohxlet's apparatus by mixtures of ethanol and dimethyl sulfoxide with their equal molar ratio (1:1). After removed of solvent from the filtrate we dried the sediment (yellow crystals) in vacuum($P_{Rem} \approx$ 0.25-0.26 kp).

Composition of synthesized dyes have been established by the element analysis (table) and by IR spectroscopic analysis. In the IR spectra absorption bands were not detected in the regions of 3390-3340 cm^{-1} which is characterized for $-NH_2$ groups. In the IR spectra the absorption bands were observed in the regions of 1650-1660 cm^{-1}, which is characterized for −N=N− groups.

It should be noted that yields of bisazodyes depend upon conditions of tetranitrogeneration stages. The research shows that with temperature increase the yields of dyes decrease due to partial destruction of intermediate tetraazonium salt and formation of corresponding bisphenol.

From obtained dyes some of them have been used for dyeing natural fibres. The dyeing was carried out in aqueous solutions by direct method. It was stated that obtained dyes at dyeing natural fibres is characterized by brightness and high fastness to different treatment.

Table

The yields, element analysis and molar mass of obtained dyes

Azoic coupling component	Formula of dyes	Yields %	Molar mass (Ebulli-oscopy)	Found % / Calculated % C	H	N	Si
I	$C_{13}O_2SiH_{26}N_4$	60	514	70.5	5.00	10.00	5.05
				70.37	5.06	10.89	5.45
II	$C_{33}O_6SiH_{26}N_4$	73.5	602	62.5	3.92	8.47	4.25
				62.78	4.32	9.30	4.65
III	$C_{33}O_6SiN_4H_{24}Na_2$	85	646	59.90	3.35	8.35	4.02
				60.30	3.72	8.67	4.33
IV	$C_{33}O_{14}SiN_4H_{30}S_4$	65	934	51.20	2.30	6.50	3.25
				50.11	3.21	6.00	3.00

I.Javakhishvili Tbilisi State University

N.Muskhelishvili Kutaisi Technical University

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ORGANIC CHEMISTRY

N.Kobakhidze, Z.Machaidze, M.Gverdtsiteli

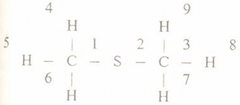
Algebraic Investigation of Thioethers

Presented by Corr. Member of the Academy L.Khananashvili, June 12, 1996

ABSTRACT. **Algebraic investigation of thioethers was carried out in terms of ANB - matrices method. Their diagonal elements represent atomic numbers of chemical elements, whereas nondiagonal ones - the multiplicities of chemical bonds.**

Contiguity matrices and their various modifications are effeciently used in modern theoretical organic chemistry [1,2]. One type of such matrices are ANB - matrices: their diagonal elements represent atomic numbers of chemical elements, whereas nondiagonal elements - the multiplicities of chemical bonds [3].

The graphic formula of dimethylsulphide (2-thiapropane) with numeration of atoms and corresponding ANB - matrices are brought below:



$$\begin{Vmatrix} 6 & 1 & 0 & 1 & 1 & 1 & 0 & 0 & 0 \\ 1 & 16 & 1 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 1 & 6 & 0 & 0 & 0 & 1 & 1 & 1 \\ 1 & 0 & 0 & 1 & 0 & 0 & 0 & 0 & 0 \\ 1 & 0 & 0 & 0 & 1 & 0 & 0 & 0 & 0 \\ 1 & 0 & 0 & 0 & 0 & 1 & 0 & 0 & 0 \\ 0 & 0 & 1 & 0 & 0 & 0 & 1 & 0 & 0 \\ 0 & 0 & 1 & 0 & 0 & 0 & 0 & 1 & 0 \\ 0 & 0 & 1 & 0 & 0 & 0 & 0 & 0 & 1 \end{Vmatrix} \qquad (1)$$

The first column of the matrix corresponds to atom numbered in graphic formula by cipher "1" (carbon-atom); second column corresponds to atom numbered by cipher "2" (sulphur-atom) etc. The first column begins with cipher "6" - the atomic number of carbon; cipher "1" means, that the bond between carbon and sulphur is simple; cipher "0" means, that carbon "1" is not bonded to carbon "3"; three ciphers "1" mean that carbon "1" and hydrogenes "4", "5" and "6" are single bonded; the column ends with three ciphers "0" - they mean that "1" carbon is not bonded with hydrogenes "7", "8" and "9". Other columns of matrix are constructed analogously.

The values of determinants of ANB - matrices do not depend on the numeration of atoms in molecules (they are invariants of molecular graphs).

For simple calculations hydrogen atoms are not often taken into account (so called "molecule skeleton" should be considered). The molecule skeleton of dimethylsulphide and corresponding simplified ANB - matrix (pseudo ANB - matrix) are brought below:

$$\overset{1}{C}-\overset{2}{S}-\overset{3}{C} \qquad \begin{Vmatrix} 3 & 1 & 0 \\ 1 & 16 & 1 \\ 0 & 1 & 3 \end{Vmatrix} \qquad (2)$$

In this case the range of ANB - matrix reduces by six units (total number of hydrogene atoms in dimethylsulphide). From diagonal elements of (1) matrix the numbers of hydrogene atoms which were bonded to corresponding atoms are substracted. It was confirmed that $\Delta(1) = \Delta(2)$.

The values of determinants of ANB - matrices for thioethers and corresponding standard entropies [4] are listed in the Table.

Table

The values of determinants (Δ) of ANB - matrices for thioethers and corresponding standard entropies (S°_{298}).

Thioether	S°_{298}	Δ
2-Thiapropane	68.32	138
2-Thiabutane	79.62	508
2-Thiapentane	88.84	1894
3-Thiapentane	87.96	1870
2-Thiahexane	98.43	7068
3-Thiahexane	98.97	6972
3,3-Dimethyl-2-Thiabutane	89.21	6264
4-Thiaheptane	107.22	25994
4-Thiaoctane	117.90	97004
5-Thianonane	125.84	361998
5-Thiadecane	136.52	1350988

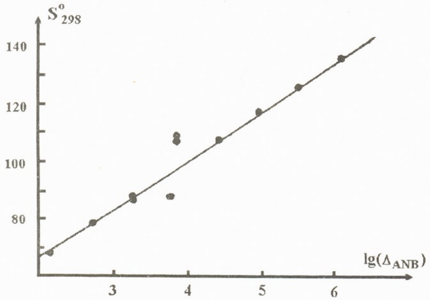


Fig.1. The plot $S^{\circ}_{298} = 17.14 \lg \Delta + 32.10$ for thioethers.

$S^{\circ}_{298} \sim \lg \Delta$ plot is constructed on the computer (Fig. 1). The equation describing this dependence has the form:

$$S^{\circ}_{298} = 17.14 \lg \Delta + 32.10 \quad (3)$$

The correlation coefficient r was calculated by formula:

$$r = \left(\sum_i y_i y_i' \right)^2 / \sum_i y_i^2 \sum_i y_i'^2 \quad (4)$$

where: y_i are values from correlation plot, y_i' - experimental values. r is equal to $r = 0.999$, so excellent correlation was observed.

Thus we can consider $\lg(\Delta_{ANB})$ as the topologic index [2] for "stucture - properties" correlation in homologous series of thioethers.

I.Javakhishvili Tbilisi State University

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №1, 1996

PHYSICAL CHEMISTRY

K.Amirkhanashvili, L.Rapoport, Academician T.Andronikashvili

Studies of Sorbent-Sorbate Adsorption Interaction on Capillary Chromatography

Presented September 2, 1996

ABSTRACT. A review is made on the capillary columns used in gas-liquid chromatographic investigation. The preference utilization of quartz capillary columns is shown.

1. Selection of a Capillary Column.

Gas-liquid chromatographic studies of components based on multiple recurrence of single acts of separation with complete phase inversion is an efficient method of investigation of physico-chemical interactions. If it is assumed that a single act of evaporation-condensation takes place on the sorbent layer within the length of one theoretical plate (Th.P.), then the investigation of physico-chemical interactions must require the conditions leading to homogeneity of Th.Ps that can be essentially characterized as:

1) minimal difference in the thickness of adsorbent layer;
2) minimal influence of the carrier on the sorption mechanism;
3) minimal difference in packing density of the sorbent lengthwise chromatography column.

While selecting the system between packed and capillary columns the pointed out requirements are complied with capillary columns thanks to the fact that a capillary wall with its small surface and limited activity serves as a sorbent carrier and the flow pipe has a large specific midship section in respect to the area occupied by the absorbent stationary liquid phase (SLPh), that makes for a relatively low fall lengthwise the column. Mentioned peculiarities quite favourable for reliable representation of physico-chemical interactions on the resulting chromatogram should make the choice unique in favour of capillary columns, but for the limitations connected with the formation of continuous, uniform and stable layer of SLPh on the column wall. As surface-tension of the layer SLPh is comparable with adhesive forces on a small radius of the capillary curvature and increases considerably for polar liquids, not all the SLPh can be applied to the capillary. They held long on its surface without disturbing continuity and formation of defects affecting the quality of the capillary columns.

Thus development of suitable technological procedures is important to overcome the mentioned obstructions.

Since columns from metal, polymers, glass and quartz have been used in capillary analytical chromatography practice [1], we should consider their suitability for performing physico-chemical studies on absorbent-sorbate systems.

Metal columns as a rule are from stainless steel or tombac, more rarely of nickel, aluminium and copper. Disadvantage of these columns is their high activity, hence metrologically non-controllable results of physico-chemical measurements. At the same time practically any SLPh can be applied to them under favourable conditions with limited but real period of column lifetime. Such columns make it possible to perform qualitative and not quantitative investigation of physico-chemical interactions.

Columns from polymeric materials are of limited use under special conditions where columns from other materials are of little use. Application of these columns in individual cases guarantee possibility of complying special requirements to the separation process on them. Columns of this type are rare, made to special order and are not on sale as a rule.

Glass columns are from plants for stretching capillaries. They are used not only for producing chromatography columns but in laboratories of educational institutions and enterprises as well wide application with gas-liquid chromatography. Application of SLPh on glass column is partially limited in respect to polarity but is quite available, though it requires considerable efforts sometimes leading to desirable reproducibility of column characteristics. Industrially made serial columns have broad nomenclature in respect to sorbents. The main disadvantage of glass columns is their fragility. Due to this disadvantage they are not applied in practice. The use of glass capillary columns in physico-chemical investigation will apparently remain limited until this disadvantage is overcome.

Quartz capillary columns were first developed by Ettre [2], who prepared and tested them in Perkin-Elmer firm. The peculiarity of such capillaries is a tube with the thickness of walls not exceeding several mkm being drawn out of quartz soft pipe of high purity and twisted in coil after cooling the melt. Thus the capillary is in stressed state, having an elliptic section differing slightly from the circular one.

To prevent germs of quartz defects growing into cracks, the outer surface of quartz is covered with polyamide coating while preparing columns. In the process of polycondensation this coating blockades the germs of defects, hindering their growth in the stressed state of the capillary walls. Besides this coating insulates the capillary from exposure to external action and gives it high flexibility.

One of the major disadvantages of such a capillary is a considerable activity of its inner surface connected with isolation of water at polycondensation of polyamide on its external surface and its diffusion to the inner surface. The only way of removing the water from the quartz wall is a blow-down of the capillary with thermal stability of the coating. Desiccation of such a capillary at an accessible temperature gives positive results only after its prolonged treatment. SLPh is applied to the inner surface by dynamic or static method [3]. The layer of SLPh may remain applied, but due to its weak adhesion to the wall, as a rule, it gains defects considerably reducing the efficiency and corrupts the nature of sorbate-absorbent interaction. In those cases, when applied layer of SLPh can be stitched, it forms immobilised layer onto the inner surface of the capillary. Its chemical energy links are enough to retain the layer as a uniform flexible formation.

Of rather wide choice of SLPh used in analytical capillary chromatography mainly methyl, methylphenyl and silicone elastomers, as well as polymethyleneglycols are amenable to immobilization. As a rule, thermal stitching has been used with peroxides of benzoyl, diphenyl or other initiators. Limitation of usable SLPh as in the general case is residual activity of the inner surface of the capillary.

P.Melikishvili Institute of Physical and Organic Chemistry
Georgian Academy of Sciences

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BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №1, 1996

GEOLOGY

Academician I.Gamkrelidze, D.Shengelia, G.Chichinadze

Macera Nappe in the Crystalline Core of the Greater Caucasus and its Geological Significance

Presented August 14, 1996

ABSTRACT. **It is indicated that in the Elbrus subzone of the Main Range zone of the Greater Caucasus are preserved separate fragments of once indivisible Macera nappe which was overthrusted from south (from the Pass subzone) during the sudetic phase. Due to this conclusion division of metamorphic complexes of the Greater Caucasus into series, their interrelations and questions of its age succession is considered in a new light.**

Since early 70-ies when North Caucasian geologists G.Baranov, I.Grekov and S.Kropachev had stated existence of Paleozoic nappes within the Forerange zone of the Greater Caucasus [1], tectonic structure of the pre-Jurassic basement of the Greater Caucasus attracted particular attention.

At the same period (1974) nappe structures were revealed in the Main Range zone of the Greater Caucasus. G.Baranov and I.Grekov and irrespectively I.Gamkrelidze and G.Dumbadze indicated that mica schists of Buulgen series are tectonically overlapped by migmatites of Macera series and the authors of this paper ascertain that the Damkhurts suite of the Laba series is tectonically overlapped by Macera series [2]. G.Baranov and S.Kropachev stated the presence of "pregranitic tectonic discontinuity" between the intra- and suprastructures of the crystalline core of the Greater Caucasus within the Northern subzone of the Main Range [1]. An idea of allochtonous nature of the Laba series and its separate parts was also expressed (G.Baranov, I.Grekov, D.Shengelia, Sh.Adamia, G.Chichinadze et al.).

Data obtained by authors during last years allow to assume, that all the stratified and nonmigmatised metamorphic suites of the Elbrus subzone represent separate fragments of once united megaslab-large nappe (Fig.1). This nappe is overthrusted on unstratified and intensively migmatized infrastructure of the Greater Caucasian crystalline core which by its primary disposition was disjoined from it. On the one hand this assumption confirms the idea that the pre-Jurassic basement of the Greater Caucasus is of imbricated structure and on the other hand allows to consider in a new light division of metamorphic complexes into series, their interrelations and age succesion.

For the overthrusted metamorphic formations we have presented the same name that was suggested by G.Baranov in 1980-"Macera nappe". Under this name he united following outcrops in the Main Range zone; between Bolshaia Laba and Mali Zelenchuk riverheads, and more to the east in Aksaut, Teberda, Baksan and Chegem riverheads.

We attribute these at present separate fragments of considerable size and also comparatively smaller ones to the Macera nappe (Fig.1). Metamorphites of all these fragments almost didn't differ in composition and degree of metamorphizm; they are

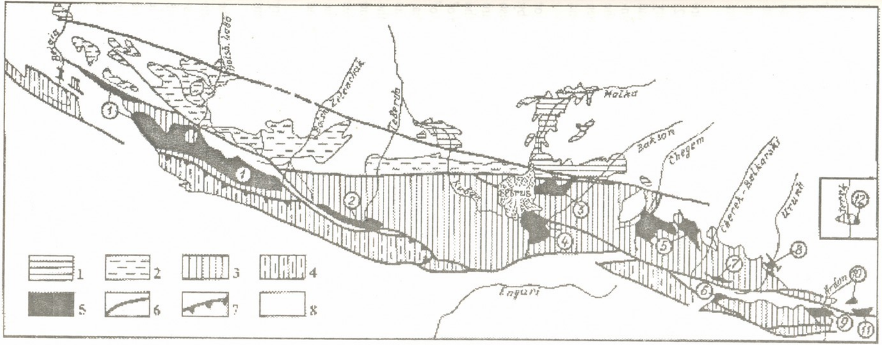

Fig.1 Scheme of disposition of Macera nappe fragments in the Main Range zone of the Greater Caucasus.
Exposures of crystalline complexes of the Greater Caucasus:
1 - in the Bechasin zone; 2 - in the Forerange zone; 3 - in the Elbrus subzone of the Main Range zone; 4 - in the Pass subzone; 5 - fragments of the Macera nappe (figures in circles): 1 - Tsakhvoa-Arkasara, 2 - Aksaut, 3 - Kirtik, 4 - Baksan (Donguzorun and Jusengi sheets), 5 - Bezengi (Koru and Siliksu sheets), 6 - Kharves, 7 - Akhsu, 8 - Matsuta, 9 - Buron, 10 - Unal, 11 - Phiagdon, 12 - Khde (Dariali); 6 - faults; 7 - exposure of Paleozoic subduction zone; 8 - Post-Paleozoic and Mesozoic nonmetamorphosed sedimentary cover.

correspondingly featured by similar mineral paragenesises and apparently in their primary disposition they occupied approximately one and the same stratigraphic level.

Geological position of separate fragments varies at present: tectonic overlap of crystalline rocks of infrastructure by metamorphities which compose them is here observed; though in some places they are izolated from each other; sometimes these fragments due to more later tectonic movements are in tectonic contact with the post-Paleozoic deposts which more often as well as Middle Carboniferous-Permian formations transgressively overlie metamorphites of separate fragments of the Macera nappe. In some places these metamorphites are tectonically layered.

In addition, in the contact zone of the Macera nappe fragments with infrastructure are stated: 1) considerable difference in the degree of their regional metamorphism; 2) development of intense cataclasis and mylonitization processes in contact roks; 3) very often disposition of post-metamorphic bimicaceous and microline granitic and granodioritic (mainly the Ulukam type) bodies in its contact area; 4) intrusions of syn- and premetamorphic granitoids are not observed in metamorphites that compose the Macers nappe; these bodies are widespread in gneiss-migmatitic infrastructures. When these bodies are in contact with the overthrusted metamorphites both granitoids and metamorphites are broken down and mylonitized and their contact is tectonic.

The Macera nappe is represented by Dombai and Arkasara (M.Somin), or Donguzorun, Dupukh, Kti-Teberda and Kurganshichat (S.Baranov) suites and also by their analogues (Bezengi, Kirtik and Buron suites). According to our data all these suites by rock and mineral paragenesises and also their metamorphic character are analogous to terrigane micaceous schists which are developed in the upper part of the Buulgen series. They or their analogues are following quite gradually the Klich suite of Buulgen series, which is underlaid by Gwandra suite. This part of metamorphites of Buulgen series is represented by nonmimatized Sisina (Ladevali, Vertskhlis tba, perhaps Uluchiran) suites. M.Somin attributes to them Dombai and Arkasara suites as well. He indicates that the Dombai suite without any break gradually continues the Klich one [3]. To our data the same picture is in interrelations of the Klich and overlying Sisina and Ladevali suites [4, 5]. According to G.Baranov the Buulgen series is terminated by Uluchiran suite which analogues he indicates in the lower part of the Macera nappe too. It's noteworthy that stratigraphic analogues of the Dombai and Arkasara suites are met both in the Elbrus and in the Pass subzone.

Thus, all the fragments of the Macera nappe corresponds to apparently once very thick upper part of the Buulgen section, and are overthrusted from south to north, from the Pass subzone into Elbrus subzone.

Taking into account that Late Hercinian granitic bodies are intruded between the infrastructure of the Greater Caucasus and Macera nappe and apparently Middle Paleozoic rocks composimg this nappe are overlapped by Middle Carboniferrous-Permian sediments, then overthrusting of the Macera nappe should be connected with Early Hercynian (Bretonic) phase, or most probably, like other Paleozoic nappes of the Greater Caucasus, with the sudetic phase, which was basically revealed here in the Lower-Middle Visean (G.Baranov et al. [9] assume the pre-Hercynian age of Macera nappe). At the same time tectonic overlap mainly on the mafic part of Buulgen series and Laba series with the infrastructure rocks of the Greater Caucasus, as it was mentioned earlier, shows that a part of these rocks at the same time are subject to underthrusting (subduction) under the island arc of the Greater Caucasus [2, 6]. (Some authors consider them an accretionary prism [7].) Besides, subduction was primary process and more prolonged in contrast to overthrusting (obduction) which had lightning speed in the geological sence [2]. We should consider that this overthrusting coinsides with the closure of marginal sea of the Southern slope of the Greater Caucasus promoted with the sudetic phase, when from the Hypsomatrically uplifted suture zone the upthrusted plate suffers gravitational sliding on the continental margin. For example, similar event took place in the process of sliding of fragments of the Patagonian back-arc basin crust in Middle Cretaceous on the continental margin [8].

Alongside with this as it was already noted, some investigators and the authors of this paper among them, consider that in recent structure Laba series is of allochtonous-imbricated composition. Here in Lashtrak, Ajara and damkhurts independent nappes are distinguished. Their primery stratigraphic succession is not clear. In first two nappes the samenamed suites are united, but in the Damkhurts nappe two suites-Damkhurts and Mamkhurts are distinguished. Analogues of the Lashtrak and Ajara suites (nappes) are known neither in stratified metamorphic complexes of the Main Range nor in Buulgen series and within the Macera nappe. In G.Baranov's opinion possible analogue of the Mamkhurts suite is Donghuzorun (Dupukh) suite but M.Somin doesn't share this opinion [3]. We think it possible that the unity Klich suite and of the certain part of micaceous schists is analogue of the Mamkhurts suite, or also Mamkhurts and Damkhurts suites, which correspond to the upper part of Gwandra suite and Klich suite. So, to our mind full synchroniety of Laba and Buulgen series and their unification into one series is inexpediently and we consider it natural, that lower parts of Buulgen series are older than Middle Paleozoic. Some authors completely attribute them to Proterozoic [9]. We can therefore conclude, that stratified members of metamorphic complex of Elbrus and also of the Pass subzone excluding Ajara and Lashtrak suites of Laba metamorphic complex before the overthrusting represented parts of indivisible thick stratigraphic section.

It follows from the all abovementioned that the old notion of "Macera series" looses its essence, as to our mind the traditional name "Macera series" unites two different geological bodies: as it seems mainly Precambrian gneiss-migmatitic infrastructural complex of Elbrus subzone of the Greater Caucasus and stratified mainly terrigene suites of Buulgen series, overthrusted from the Pass subzone which may be of Middle Paleozoic age.

In order to define tectonic position of metamorphic complexes of the Greater Caucasian crystalline core it is essential to make comparative petrological characteristic of granitoids which had developed within the limits of these complexes. It is quite natural that granitoids which developed in different structural zones, and hence, in geological retrospective in different geodynamic conditions and in the basement of various composition, essentially differ from each other.

In conditions of selective melting of sialic infrastructure of Elbrus subzone (island arc of the Greater Caucasus) which took place immediately after overthrusting of Macera nappe, in conditions of deep subsidence of this subzone were formed K-rich granites; their considerable part emplacemented in the contract zone of the infrastructure and the Macera nappe and "heal" their tectonic contact. It's quite natural that these granites cut syn- and premetamorphic granitoids which are widespread in the infrastructure and their pebble are observed in Upper Carboniferrous and Permian sediments.

In the same time interval (sudetic phase) in the Pass subzone in lower part of Buulgen series and mainly in underlying, deeper horizons magmatites of low-K quartz-diorite-plagiogranitic series are formed. Through the whole section of Buulgen series diorite-plagiogranitic series are formed. Through the whole section of Buulgen series their active contact whith metamorphites is observed. Varieties of low-K postmetamorphic granitoids are not found within the Macera nappe and post metamorphic high-K granites whith form in infrastructure are not found in metamorphits of Buulgen series.

As for syn- and premetamorphic granitoids in Elbrus zone they exist in the infrastructure. Synmetamorphic granitoids are represented by Early Hercynian plagiogneisses, plagiogranites, plagiomigmatites, microclinised porphyroblastic granites. Their pebble appear in Middle Carboniferous sediments. Orthogneisses of granodioritic composition belong to premetamorphic (pre-Hercynian) granitoids. In the Pass subzone the premetamorphic granitoids (gneissic diorites, quartz diorits, plagiogranites, plagiogranodiorites) occur in the lower and middle parts of the Buulgen series section, and Synmetamorphic granitoids are represented by various para- and orthogneisses and granitoids.

In Laba metamorphic complex are notable premetamorphic, apparently Lower Paleozoic Beshta and Kamenistaia plagiogranodioritgneisses, whose analogues are unknown in the Main Range zone. They are located within the Lashtrak and Ajara nappes [5]. Whereas to synmetamorphic granitoids belong plagiogneissic and plagiogranitic sheets of the Mamkhurts suite and gneissic quartz dioritic and autochthonous conformable bodies of plagiogranites of the Damkhurts suite.

Finally, if we look through the present day data on tectonic structure of the Greater Caucasian pre-Jurassic basement, we should share the idea that it represents a typical tectonically layered complex [10]. Such horizontal tectonic layering of pre-Jurassic basement not only in the Greater Caucasus, but also through the whole Caucasus and its adjacent regions was indicated before [6]. At the same time its striking that most of the overthrusted slabs of the Greater Caucasus were formed in different geodynamic conditions, particullarly, here are combined formational complex of palezoic island arc and marginal sea and various parts of oceanic crust of Paleotethys basin (in Forerange zone). If we also take into account, that most of the overthrusted slabs were moving on the serpentinite "lubricant" (in their contact zones bodies of ultramafites are very frequent), we should consider, that here is real vertical-accretionary complex which consists of separate fragments of exotic terranes (classic example of such accretion is apparently Brooks Range in the northern part of Alaska). In other words, during the Hercynian tectogenesis, and may be partially during Early Cimmerian folding took place mainly vertical and not lateral accretion of the earth crust; that naturally complicates to define primary (predeformational) spatial disposition of separate terranes. This problem requires special analyses and is not a subject of this paper.

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
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H.Mikadze

Method of Compact Limestones Treatment for Micropaleontological Investigation

Presented by Academician I.Gamkrelidze, June 06, 1996

ABSTRACT. **A new method is applied for separation of microforaminifers from the Upper Cretaceous compact limestones, glauconitic sandstones and clay and marlaceous limestones of the Gagra-Djava zone.**

Upper Cretaceous deposits of the Gagra-Djava zone are represented by carbonaceous rocks; variegated limestones containing frequent nebses and intercalation of variegated flints.

Many investigations suggested various zonal subdivisions of the Upper Cretaceous sediments of the region. In most cases foraminifers were studied either in petrographic sections or in the samples from the soft rocks. But in petrographic sections precise identification of taxons is impossible as the shell are of diverse orientation and the section crossouts them in different planes.

Thus resulting from the above mentioned it is indispensable to work out a new method of foraminifer separation from compact rocks. Collection of foraminifers from the Upper Cretaceous (Abkhazis-Racha and Odishi-Okriba) facies type sediments are used for treatment. Samples were taken in the proportions from 30 cm up to 2 m and more, according to the character of rocks; 1000 samples alltogether.

Techniques of Sample Treatment

Fig.1. Marginotruncana pseudolinneiana Pessagro, Georgia (Samegrelo), the Tskhemura river (left tributary of the river Tekhura), Lower Turonian

For the first time in Georgia we have applied threestaged method of sample treatment allowing to desintegrate compact rocks and to separate foraminifer shells.

Main point of the method: on its first stage 200-300 g of the rocks is pestled in the iron mortal; size of fractions shouldn't exceed 0.1-0.5 mg. Solution of sodium carbonate is poured on the obtained powder 1 tablespoon full of sodium carbonate ($Na_2CO_3 10{\cdot}H_2O$) to 250 g of water.

We keep this powder in solution during 24 hours, then we heat it until its evaporation and exsiccation.

The second stage is treatment of the powder in the hydrous sodium sulphate. Hydrous sodium sulphate is heated beforehand. During calcination it extracts water and starts boiling. In the process of boiling the powder produced at the first stage is added. This porridge like mass is heated and cooled for several times (5-10 times) then it is washed in water and dried. It is washed off in metal or in some other fireproof vessel.

Fig.2. Globotruncana bulloides Volger, Georgia (Samegrelo), the Tskhemura river, (left tributary of the Tekhura river), Lower Santonian.

At the third stage rocks are treated with glacial acetic acid and copper vitriol. Copper vitriol (5-10 g) on each sample is treated with fire (only enamel vessels are used), till it becomes white hot and water molecules completely evaporate.

On the second stage preliminarily washed and dried mass is put into the tin. Worked up copper vitriol is added here and afterwards 20-25 mg of glacial acetic acid is added and the tin is tightly closed.

As we know the glacial acetic acid freezes under +17°C. During 3-4 days this solution is alternately kept in comparatively cool (17°C) and warm (17°C) place. After that the powder is washed out in water, till it becomes transparent.

After each stage obtained mass is washed out in warm water.

As we know carbonate sediments as well as shells of foraminifers are dissoluted under the influence of acid. In order to choose the most optimal time for keeping the powder in the glacial acetic acid and cooper vitriol without distroying the shells several tests are carried out.

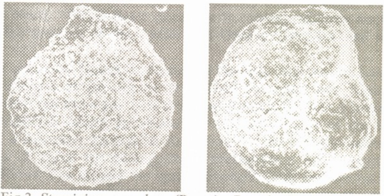

Fig.3. *Stensioina oxsculpta* (Reuss), Georgia (Samtredia) the Tskhemura river, (left tributary of the Tekhura river), Upper Coniacian
a - dorsal view
b - ventral view

The most suitable sample is pestlled and put into several tins (the weight of powder doesn't matter) .Then we repeat the third stage of washing out. But time interval is different for each portion of the powder. First portion is washed out after 12 hours; the second one - after 24 hours, the third - after 2 days and the fourth - only after 4 days. After each washing out the powder is studied with ocular and after these tests it turned out that the optimum time for keeping the powder in acid without distroying the shells is more than 4 days for very compact rocks and 1-2 days for the soft ones.

Basing on the above mentioned method used on the samples taken from several sections of the Gagra-Djava zone we can make the following conclusions:

1. For glauconitic sandstones that are mainly observed in the Cenomanian, first stage of the method must be used.

2. For the clay and marlaceous limestones that don't comprise flinty intercalations - first and second stages are forbidden, as water and hydrous sodium sulphate will distroy the shells of foraminiferes (shells and rocks are of the same chemical composition). Thus only the third stage of the method can be used here.

3. All three stages can be applied to compact flinty limestones. It should be noted that after the second stage of washing, heary concentrate must be studied under the binocular for sampling of those foraminifers which do not require further processing. The rest of the material is treated according to the third stage.

Thus the new method of washing out of the heavy concentrate allows to distinguish well preserved abundant planctonic and benthonic foraminifers.

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
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PETROLOGY

A.Okrostsvaridze, D.Shengelia

New Data on the Formation of Microcline Granites of Dzirula Salient

Presented by Corr. Member of the Academy G.Zaridze July 8, 1996

ABSTRACT. **Considering structure of Dzirula salient characteristics of binary granites formation of inner Massif of the Caucasus it is assumed that during Tournasion-early-middle Visean period the rocks of sialic profile (the substratum of microcline granitoids) were overthrust by allochthons of femic profile. The thickening of the crust caused the change in PT gradient which stimulated the process of selective melting in sialic rocks and formation of S-type granites. The trace of tectonic transport of allochton was closed and thus erased by granitoid magma activity (late Visean-Bashkirian) in major areas.**

The Caucasus is spread from the Caspian Sea to the Black Sea and represents a connective segment between the European and Asian parts of the Alpine-Himalayan folded system. It is situated between comparatively static Russian and Arabic plates, moving to the North - North-West direction. It consists of three major geostructural units: the folded systems of the Lesser and Greater Caucasus and inner Caucasion Massif.

The inner Caucasion Massif is situated in the area between the Lesser and Greater Caucasus. It consists of pre-Alpine crystalline basement and Alpine Sedimentary cover. In the latter one there are three principal salients: Dzirula, Khrami and Loki.

Dzirula crystalline salient is situated in the central part of the inner Caucasian massif and is a complex geological formation. In the Dzirula salient two-age and genetic groups of rocks are distinguished: prehercynian and hercynian formations. Prehercynian formations occupy the upper structural floor and are mainly represented by crystalline shales, plagiogneisses, amphibolites, quartz-diorite gneisses. In this association of rocks dominates the last one. Numerous inclusions of different shape and size of gabbroes, gabbro-diorites and diorites are observed in them with diameters in the range of 3-60 cm and in some places comprise the half of the outcropped area. In tectonic scales of the farther east of Dzirula salient assosiation of serentinites gabbro-amphibolites and diabase porphyrites are distinguished. They are considered as ophiolitic formations [1,2]. It should be mentioned that in 1953 G.Zaridze and N.Tatrishvili considered femic rocks of Dzirula salient as residues of Paleozoic geosynclinal substratum and that quarz-diorite gneisses were formed as a result of metasomatic granitization [3]. This idea was approved in geological circles and in fact it gave start to the principles of metasomatic granitization. I.Khmaladze and K.Chikelidze [4] considered the above mentioned quartz-diorite-gneisses as magmatic formations explained by hypodiomorphic structures, zonal plagioclasses and unchangable mineral and chemical composition and disoriented xenolithes. According to our data quarz-diorite gneisses of Dzirula salient are hybrid formations formed under the influence of magmatic material on progressive sedimentary sheet by the so-called melting. According to Rb/Sr and U/Pb dating the absolute age of quartz-diorite gneisses of Dzirula salient is evaluated as 762±222 Ma [5].

The rocks described are crossed by microcline granites, which form sectioning bodies of different thickness. In fact all the old formations are full of the bodies of such veins and injections: as a result crystallin plates plagiogneisses, plagiomigmatites, tonolites and metabasites undergo microclinization of different quality. In fact microclinization gives the ultimate geological picture of Dzirula salient. The sectioning bodies are represented by anatectic massive equally-grained biotite-muscovite granites, alaskites, aplites and pegmatoids. The main rock-constructing minerals are quartz, acid plagioclase, microcline biotite and muscovite. According to Rb/Sr and U/Pb the absolute age of binary granites of Dzirula salient is 330-311 Ma [5].

Petrochemical study of quartz-diorites gneisses and microcline granites and also the data of discriminant diagram by Hassan and McAlister [6] show (Fig.1) that quartz-diorites of hybrid genesis are attributed to granitoids of I-type while anatectic Microcline-granites represents granitoids of S-type.

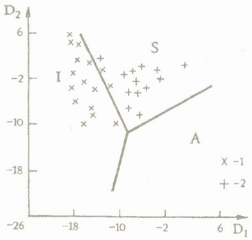

Fig. 1. Prejurassic granitoids of Dzirula salient on the discriminant I, S, A diagram [6].

$D_1=0,76Al_2O_3+5,96TiO_2+2,91MnO-1,93Na_2O+1,95K_2O-18,50P_2O_5$

$D_2=0,37\ Al_2O_3+7,25TiO_2-54,04MnO-4,28Na_2O-0,55K_2O+45,81P_2O_5$

1. Quartz-diorite gneisses, 2. Microcline Granitoids

As we can see quartz-diorite gneisses and microcline granites are separated not only in time but they are genetical disconnected as well. The first ones are prehercynian hybrid formations formed by complex interaction between the mantle and crust by the so called "progressive melting" and the others are formed in the late Hercynian period by typical anatectic melting of sialic material.

Thus the rocks of femic complex of Dzirula salient occupying, the upper floor of the structural section can be considered as the residue of the oceanic crust as a result of tectonic and metasomatic processes; here ophiolitic scales are distinguished, but the formations of gabboro-diorite-quartzdiorite series prevail with fragments of regionally metamorphosed sedimintary and volcanic sedimentary rocks.

But microcline-granites that occupy the lower floor of the section are apparently the result of continental type sediments melting.

It should be mentioned that none of the concerning works can explain the existing picture of formation of granitoids. According to the present view two-mica potassium granitoids should be formed as a result of selective melting of poor in potassium gabboro-diotite-quartz diorate series, or of deeper basalts and ultrabasits of the oceanic crust; this is impossible even by the results of experimental data. For real picture the only scheme is to admit the reverse variant of the construction of the crust of Dzirula salient of the Caucasus inner Massif. Here the upper sections of searm outcrop is occupied by the oceanic crust formation and lower - by the continental crust. Following from this we admit that the rocks of salic profile (granitoid substratum) were covered by a thick overthrust - sheet of femic profile.

The period of formation of charriage is the same as of Chorchano-Utslevi allochthon complex: allochthon complex: Tournasion-early-middle-Viseon, because, according to all features they are the result of one and the same tectonic process. The thickness of allochton reached several kilometers and thickening of the crust established high PT gradient and stimulated in sialic rocks the process of selective melting and the formation of two-mica granites. Their formation took place already in late-Visean-Bashkirion period, because they have intrusive contacts with Chorchano-Utslevi allochthon, but do not cross upper carbonate, neoautochton. It should be mentioned that activity of granite magmo healed and erased the zone of allochthon transport.

Besides the above mentioned data the following facts indicate allochthonous character of the femic complex of Dzirula salient bulge:

1. Metabasites analogical to intensively tectonized ophiolites (the second and third layers of the oceanic crust) or paleooceanic crust of middle oceanic range type are widely spread in them [7].
2. Close space and genetic relation of the rocks of femic complex with the allochthonous scales of Chorchano-Utslevi.
3. The different structure and style of the rocks of femic and sialic profil [2].

Thus we have an interesting picture of formation of microcline granites in Dzirula salient of inner Massif of the Caucasus. Namely during Tournasion-early-middle-Visean period formation of chariage of thick femic rocks (the crust of oceanic type) over the rocks of sialic profile (the crust of continental type) took place, which caused thickening of the earth crust in this area and estableshed high PT gradient. This process stimulated selective melting in sialic rocks and formation of microcline S-type granites, which during late Visen-Bashkirian period crossed and metasomatically changed the upper layer of the crust of oceanic type.

A.Janelidze Institute of Geology
Georgian Academy of Sciences

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METALLURGY

A.Sulamanidze

The Way of Boundary Effect Detection at Conductor Setting

Presented by Academician T.Loladze, June 27, 1996

ABSTRACT. **An original way of detection of thermoelectromotive force originated on the border of solid and liquid phases of current conducting material is proposed. A simple and accessible scheme is worked out for registration of the mentioned effect.**

Practical application of the method opens possibilities of quantitative estimation of boundary effect of conductors, fixing their place in thermoelectric metal series and of consecutive control of thermal processes in the zone of unipolar current pulse resistance welding and of crystallization process control.

As far back as 1956 A.Ioffe [1] expressed an idea and then in 1957 W.Pfann, K.Benson and J.Wernick [2] proved experimentally that thermoelectric phenomenon initiated on the border of solid and liquid states of the same conductor must have a definite effect on metal recrystallization process. Namely, heat removal emitted at this time occurs usually not only with heat conduction, but with thermoelectric effect electron flow as well. Therefore the mentioned thermoelectric effect can be used as means of thermal processes regulation in the zone of unipolar current pulse of electric resistance welding and also for crystallization process control.

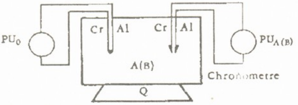
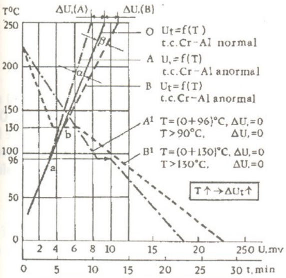


Fig. 1. Practical scheme of the detection effect.

For practical realization of the latter notion and considering that nowadays thermoelectric process occurring on the border of metal phase transformation is not almost investigated, the author has developed an original scheme (Fig. 1). It suggests placement of the investigated material, the alloys A and B, in fireproof porcelain melting-pot. The chemical composition of these materials are shown in Fig. 2, Then the material is heated with external heat source (Q) above melting temperature for (100-150°C). Chromel-alumel (two) thermocouples, one of which (left) is normal and the second (right) has shared ends are placed in melted material. The readings of thermocouples are simultaneously registered with two millivoltmeters. The left one (PU) registers the readings of normal thermocouple and the right one (PU) the readings of the abnormal one.

Investigation process is carried out first for one (A) and then for the second (B) material. When the temperature of melted material becomes 250^0C, chronometer is switched on and measurement results (temperature, voltage fall) are in equal time intervals fixed in tables. Chromel-alumel thermocouple characteristics (straight line) known from literature is at first marked down on the corresponding reference axes prepared on the basis of table data (Fig.2). Then diagram (A) is plotted for material A using the data obtained with shared ends thermocouple (millivoltermeter PU) and for the same material A, the setting diagram (A) is plotted basing on chronometric data. Thus for material A we have three graphs (O, A, A), wherefrom it is evident that when it transfers from liquid to solid state to setting, the data of two millivoltmeters are intershifted for a definite angle (α) (called below as effect characterizing angle) and their intercoincidence begins from the point which corresponds to setting temperature (96^{0}C) of this material.



n\c	Sn	Pb	Cd	Bi	Zn	In
A	18,75	31,25	—	50,0	—	—
B	52,0	30,0	13,0	5,0	—	—

Fig. 2. Diagrams of material setting and thermoelectric characteristics of the effect.

Georgian Technical University

Quantitatively different, but qualitatively analogous picture is observed for material B - three graphs (O, B, B) and characteristic angle (β). Intercoincidence of (O) and (B) graphs and the existence of divergence angle (β) ceases when melting temperature of material reduced to setting temperature (130°C).

The obtained results prove the considerations expressed by the author as a result of his recent works [3, 4].

It should be mentioned that the so-called effect characterizing angle detected by this method allows to register the border effect of a conductor, quantitatively estimate and to compose thermoelectric series of conductors and semiconductors according to potential gradients that occur on the border of solid and liquid phases of separate materials.

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HYDRAULIC ENGINEERING

D.Namgaladze, V.Nanitashvili

Nonstationary Process Control in Supply Pipelines by the Finite Control Method

Presented by Academician G.Chogovadze, June 22, 1996.

ABSTRACT. **In the paper problem of pressure control at the beginning and at the end of supply pipeline is considered. The method of finite control is used. On the basis of differential equations describing the nonstationary operation regime in the pipeline and using optimal control theory, the control functions are obtained. They allow to reduce pressure oscillations to a minimum in the pipeline at transfer from one stationary regime to another.**

A supply pipeline is a complex engineer construction. Reliable operation of a pipeline depends on a number of factors and is strongly connected with strength of largediameter pipes operating under high pressures.

At present strength calculation in supply pipelines is performed by the limiting state method considering tensile strength under the static internal pressure. This calculation does not allow possible inhomogeneity of stress distribution on the pipe wall caused by deviation of the pipe profile from the right geometrical shape due to the bead, its edge shift and profile ovality, as a whole.

In practice there is a certain failure percentage of pipes, satisfying static strength conditions while they wear.

The most probable cause of these failures is accumulation of damages and development of initial defects resulting in appearance and spreading of cracks due to repeated effects in the process of operation. Thus according to [1], supply pipeline undergo approximately 500÷600 cycles of repeated loadings a year caused by various technological and operation factors (pumping station switching off, caused by failures in electric and mechanical equipment, changes in pumping regime, etc.).For the estimated service period of a pipeline (≈20 years) total number of loading cycles can excess over 7000÷9000 cycles.

Let us consider the transfer of the supply pipeline from one stationary regime to another. In this case pressure and discharge fluctuations occur. Therefore such pressure variation with time at the pumping station should be chosen, when these fluctuations might be removed in the shortest possible time. The optimal control theory offers several ways to solve this problem, in particular, the finite control method [2] which is developed for a specific problem [3].

Let us consider pressure distribution in the pipeline, with respect to the new stationary state, as a perturbation occuring instantly at the moment of change of the pipeline operation regime.

The non - stationary regime of the pipeline operation is described by a linearized set of I.A.Charnyi equations[4]:

$$\left.\begin{aligned} -\frac{\partial P}{\partial x} &= \frac{\partial v}{\partial t} + 2av; \quad 0 \le X \le L \\ -\frac{\partial P}{\partial t} &= C^2 \frac{\partial v}{\partial x}; \qquad t \ge 0 \end{aligned}\right\}, \tag{1}$$

where P and v - the pressure and mass velocity, respectively;
c - the velocity of shock wave propagation;
a - the linearization constant.
Initial conditions are:

$$\left.\begin{aligned} P(x,0) &= P_H(x) = P_H + \frac{x}{L}(P_0 - P_H) \\ \frac{\partial P}{\partial t}(x,0) &= 0 \end{aligned}\right\}, \tag{2}$$

where P_H - the pressure at the beginning of the run in the initial stationary regime;
P_0 - the pressure at the end of the run in the initial stationary regime.

After a new stationary regime has been established, pressure distribution in the pipeline is described by the expression:

$$P_k(X) = P_k + \frac{x}{L}(P_0' - P_k), \tag{3}$$

where P_k and P_0' - pressures at the beginning and at the end of the pipeline, respectively.

Such situation can be observed when an additional pump is connected to the line.

Boundary conditions are chosen as follows:

$$\left.\begin{aligned} P(0,t) &= P_k + U(t) \\ P(L,t) &= P_0' + F(t) \end{aligned}\right\} \tag{4}$$

where $U(t)$ and $F(t)$ are the control functions chosen so that for a finite time T a new stationary regime might be established in the pipeline.

The stationary condition

$$\frac{\partial P(x,T)}{\partial t} = 0 \tag{5}$$

is added to the condition (3).

From Equation (1) one can easily obtain:

$$\frac{\partial^2 P}{\partial t^2} + 2a\frac{\partial P}{\partial t} - c^2\frac{\partial^2 P}{\partial x^2} = 0 \tag{6}$$

If a new function $R(x,t) = P(x,t) - P_k(x)$ is introduced it will be evident that

$$\frac{\partial^2 R}{\partial t^2} + 2a\frac{\partial R}{\partial t} - c^2\frac{\partial^2 R}{\partial x^2} = 0 \tag{7}$$

If we denote $\Delta p = P_H - P_k$; $\Delta p' = P_0' - P_0$; $\Delta p'' = \Delta p - \Delta p'$, it can be obtained:

$$R(x,0) = \Delta p''\left[1 - \frac{x}{L}\right] - \Delta p'. \tag{8}$$

New condition take the form

$$\left.\begin{array}{ll} \dfrac{\partial R(x,0)}{\partial t}=0; & R(0,t)=U(t); \\ R(L,t)=F(t); & R(x,T)=0; \\ \dfrac{\partial R(x,T)}{\partial t}=0; & \beta=\dfrac{aL}{\pi c}. \end{array}\right\}. \tag{9}$$

Physically minimum time for solution of the set problem is $T = 2L/c$.

Let us introduce dimensionless coordinates $\xi = \pi x/L$ and $\tau=\frac{\pi ct}{L}-\pi.$, as well as a new function $Z(\xi,\tau)=\frac{R(x,t)+\Delta p'}{\Delta p''}$. Then we obtain the following problem:

$$\frac{\partial^2 Z}{\partial \tau^2}+2\beta\frac{\partial Z}{\partial \tau}-\frac{\partial^2 Z}{\partial \xi^2}=0 \tag{10}$$

$$\left.\begin{array}{ll} Z(\xi,-\pi)=1-\dfrac{\xi}{\pi}; & \dfrac{\partial Z(\xi,-\pi)}{\partial \tau}=0; \\ Z(0,\tau)=\dfrac{U(t)+\Delta p'}{\Delta p''}; & Z(\pi,\tau)=\dfrac{F(t)+\Delta p'}{\Delta p''}; \\ Z(x,\pi)=\dfrac{\Delta p'}{\Delta p''}; & \dfrac{\partial Z(x,\pi)}{\partial t}=0. \end{array}\right\} \tag{11}$$

Hereafter we shall write x and t instead of ξ and τ.

Let us denote

$$(t)=\frac{U(t)+\Delta p'}{\Delta p''} \quad \text{and} \quad M(t)=\frac{F(t)+\Delta p'}{\Delta p''} \tag{12}$$

Introduce a new function $S(x,t)$ instead of $Z(x,t)$

$$Z(x,t) = S(x,t)e^{-\beta(t+\pi)}. \tag{13}$$

Then our problem will be changed into the following one:

$$\frac{\partial^2 S}{\partial t^2}-\beta^2 S(x,t)-\frac{\partial^2 S}{\partial x^2}=0\,, \tag{14}$$

$$\left.\begin{array}{l} S(0,t)=Z(0,t)e^{\beta(t+\pi)}=V(t)e^{\beta(t+\pi)} \\ S(\pi,t)=Z(\pi,t)e^{\beta(t+\pi)}=M(t)e^{\beta(t+\pi)} \end{array}\right\} \tag{15}$$

Let us denote

$$\left.\begin{array}{l} r(t)=V(t)e^{\beta(t+\pi)} \\ z(t)=M(t)e^{\beta(t+\pi)} \end{array}\right\} \tag{16}$$

Since $\tau=\dfrac{\pi ct}{L}-\pi$, at $t = 0$, $\tau = -\pi$, i.e. at $\tau < -\pi$ or $t < -\pi$ the problem becomes meaningless. Therefore with $t < -\pi$, $S(x, t) = 0$.

That is at $t < -\pi$ the function $S(x,t)$ can be continued with zero. Therefore equation (14) will be written down as follows:

$$\frac{\partial^2 S}{\partial t^2} - \beta^2 S - \frac{\partial^2 S}{\partial x^2} = \left(1-\frac{x}{\pi}\right)\left[\beta\delta(t+\pi) - \frac{d}{dt}\delta(t+\pi)\right] \tag{17}$$

Let us make Fourier transformatin of equation (17) over t variable; we obtain

$$-\omega^2\widetilde{S}(x,\omega) - \beta^2\widetilde{S}(x,\omega) - \frac{d^2\widetilde{S}(x,\omega)}{dx^2} = \left(1-\frac{x}{\pi}\right)e^{i\omega\pi}(\beta+i\omega) \tag{18}$$

The boundary conditions:

$$\left.\begin{aligned} \widetilde{S}(0,\omega) &= \widetilde{r}(\omega) \\ \widetilde{S}(\pi,\omega) &= \widetilde{\tau}(\omega) \end{aligned}\right\} \tag{19}$$

The solution of equation (18)can be written as

$$\widetilde{S}(x,\omega) = c_1 h_1(x,\omega) + c_2 h_2(x,\omega) + \widetilde{F}(x,\omega), \tag{20}$$

where $h_1(x,\omega) = e^{i\sqrt{\omega^2+\beta^2}x}$ and $h_2 = e^{-i\sqrt{\omega^2+\beta^2}x}$.

$$F(x,\omega) = -\sum_{k=1}^{2} h_k(x,\omega)\int_0^x \frac{W_k[h_1(\xi,\omega),h_2(\xi,\omega)]}{W[h_1(\xi,\omega),h_2(\xi,\omega)]}(\beta+i\omega)e^{i\omega\pi}\left(1-\frac{\xi}{\pi}\right)d\xi.$$

Here $W[h_1,h_2]$ is Wronsky determinant; $W_k[h_1,h_2]$ are the determinants obtained from the Wronsky determinant by replacing the k - column with the column (0;1). After calculating W_k determinants and making integrations we obtain

$$\widetilde{F}(x,\omega) = \frac{1}{\lambda^2}(\beta+i\omega)e^{i\omega\pi}\left[\frac{x}{\pi} - 1 + \cos\lambda x - \frac{\sin\lambda x}{\pi\lambda}\right] \tag{21}$$

where $\lambda = \sqrt{\omega^2+\beta^2}$.

Integration constants c_1 and c_2 should be chosen so, that the boundary conditions might be fulfilled. Therefore we obtain:

$$\left.\begin{aligned} c_1 h_1(0,\omega) + c_2 h_2(0,\omega) + \widetilde{F}(0,\omega) &= \widetilde{r}(\omega) \\ c_1 h_1(\pi,\omega) + c_2 h_2(\pi,\omega) + \widetilde{F}(\pi,\omega) &= \widetilde{\tau}(\omega) \end{aligned}\right\} \tag{22}$$

Solving this system, we obtain:

$$\widetilde{S}(x,\omega) = \frac{1}{\Delta(\omega)}\left[h_1(x,\omega)\Delta_1(\omega) + h_2(x,\omega)\Delta_2(\omega)\right] + \widetilde{F}(x,\omega), \tag{23}$$

where

$\Delta(\omega) = -2i\sin\lambda\pi$

$$\left.\begin{aligned} \Delta_1(\omega) &= \widetilde{r}(\omega)e^{-i\lambda\pi} - \widetilde{\tau}(\omega) + \frac{(\beta+i\omega)e^{i\omega\pi}}{\lambda^2}\left[\cos\lambda\pi - \frac{\sin\lambda\pi}{\lambda\pi}\right] \\ \Delta_2(\omega) &= \widetilde{\tau}(\omega) - \frac{(\beta+i\omega)e^{i\omega\pi}}{\lambda^2}\left[\cos\lambda\pi - \frac{\sin\lambda\pi}{\lambda\pi}\right] - \widetilde{r}(\omega)e^{i\lambda\pi} \end{aligned}\right\} \tag{24}$$

Since the $S(x,t)$ function must be finite over the t variable (different from idential zero at $|t| < \pi$), according Vinier-Pely theorem its Fourier representation $\widetilde{S}(x,\omega)$ at $x \in [0;\pi]$ must satisfy three requirements:

1. $\widetilde{S}(x,\omega)$ can be analytically continued over the whole complex plane $z = \omega + iy$ (i.e. $\widetilde{S}(x, z)$ function must be an integral function over the complex variable z);
2. $\widetilde{S}(x, z)$ must have the degree no more than π;

3. $\widetilde{S}(x,\omega)$ must be integrable with the square on the real axis $z = \omega$.

To fulfill the requiment it is necessary and sufficient that $\Delta_1(z_k) = 0$ or $\Delta_2(z_k) = 0$ at all points, where $\Delta(z_k) = 0$. Thus from equation (23) we obtain $\sin\sqrt{z_k^2+\beta^2}\pi = 0$ or $z_k = \pm\sqrt{k^2-\beta^2}$.

Theorefore the first requiment takes the form of the following interpolation problem for the functions:

$$e^{-i\lambda\pi}\widetilde{r}(\omega) - \widetilde{\varkappa}(\omega) = \frac{(\beta+i\omega)e^{i\omega\pi}}{\lambda^2}\left[\frac{\sin\lambda\pi}{\lambda\pi} - \cos\lambda\pi\right] \tag{25}$$

$$e^{i\lambda\pi}\widetilde{r}(\omega) - \widetilde{\varkappa}(\omega) = \frac{(\beta+i\omega)e^{i\omega\pi}}{\lambda^2}\left[\frac{\sin\lambda\pi}{\lambda\pi} - \cos\lambda\pi\right] \tag{26}$$

or

$$(-1)^k\widetilde{r}\left(\pm\sqrt{k^2-\beta^2}\right) - \widetilde{\varkappa}\left(\pm\sqrt{k^2-\beta^2}\right) = -(-1)^k\frac{\beta \pm i\sqrt{k^2-\beta^2}}{k^2}e^{\pm i\sqrt{k^2-\beta^2}\pi}. \tag{27}$$

Let us write equation (27) for even and odd k and obtain

$$\widetilde{r}\left(\pm\sqrt{k^2-\beta^2}\right) - \widetilde{\varkappa}\left(\pm\sqrt{k^2-\beta^2}\right) = -\frac{\beta \pm i\sqrt{k^2-\beta^2}}{k^2}e^{\pm i\sqrt{k^2-\beta^2}\pi}, \tag{28}$$

$$-\widetilde{r}\left(\pm\sqrt{k^2-\beta^2}\right) - \widetilde{\varkappa}\left(\pm\sqrt{k^2-\beta^2}\right) = \frac{\beta \pm i\sqrt{k^2-\beta^2}}{k^2}e^{\pm i\sqrt{k^2-\beta^2}\pi}. \tag{29}$$

To find functions $\widetilde{r}(z) - \widetilde{\varkappa}(z)$ and $\widetilde{r}(z) + \widetilde{\varkappa}(z)$ use the interporation Lagrange formula ($\sum_k{}'$ and $\sum_k{}''$ symbols denote summation over even and odd k).

$$\widetilde{r}(z) - \widetilde{\varkappa}(z) = -\sum_k{}'\left[\frac{\beta+i\lambda_k}{k^2}e^{i\lambda_k\pi}\frac{\widetilde{\varphi}(z)}{\widetilde{\varphi}'(\lambda_k)(z-\lambda_k)} + \frac{\beta-i\lambda_k}{k^2}e^{-i\lambda_k\pi}\frac{\widetilde{\varphi}(z)}{\widetilde{\varphi}'(-\lambda_k)(z+\lambda_k)}\right] + $$

$$+\widetilde{\gamma}(z)\widetilde{\varphi}(z), \tag{30}$$

$$\widetilde{r}(z) + \widetilde{\varkappa}(z) = -\sum_k{}''\left[\frac{\beta+i\lambda_k}{k^2}e^{i\lambda_k\pi}\frac{\widetilde{\varphi}(z)}{\widetilde{\varphi}'(\lambda_k)(z-\lambda_k)} + \frac{\beta-i\lambda_k}{k^2}e^{-i\lambda_k\pi}\frac{\widetilde{\varphi}(z)}{\widetilde{\varphi}'(-\lambda_k)(z+\lambda_k)}\right] + $$

$$+\widetilde{\gamma}(z)\widetilde{\varphi}(z), \tag{31}$$

where $\lambda_k = \sqrt{k^2-\beta^2}$,

$\widetilde{\varphi}(z)$ - the integral function of z argument,

$\widetilde{\varphi}(\pm\lambda_k) = 0$,

$\widetilde{\varphi}'(\pm\lambda_k) \neq 0$,

$\gamma(z)$ - the integral function of zero degree.

For satisfaction of other requirements of the Vinier – Pely theorem and for the functions to be finite on the segment $[-\pi, +\pi]$, it is sufficient to take the function $\varphi(z)$

$$\tilde{\varphi}(z)=\frac{\sin\sqrt{z^2+\beta^2}\pi}{\sqrt{z^2+\beta^2}} \tag{32}$$

then

$$\tilde{\varphi}'(\pm\lambda_k)=\frac{\pm\pi\lambda_k(-1)^k}{k^2} \tag{33}$$

With allowance for equations (32), (33) and equations (30), (31) takes the form:

$$\tilde{r}(z)-\tilde{\pi}(z)=-\left[\sum_k{}'\left(\frac{\beta+i\lambda_k}{z-\lambda_k}\frac{e^{i\lambda_k\pi}}{\pi\lambda_k}-\frac{\beta-i\lambda_k}{z+\lambda_k}\frac{e^{-i\lambda_k\pi}}{\pi\lambda_k}\right)+\tilde{\gamma}(z)\right]\tilde{\varphi}(z), \tag{34}$$

$$\tilde{r}(z)+\tilde{\pi}(z)=\left[\sum_k{}''\left(\frac{\beta+i\lambda_k}{z-\lambda_k}\frac{e^{i\lambda_k\pi}}{\pi\lambda_k}-\frac{\beta-i\lambda_k}{z+\lambda_k}\frac{e^{-i\lambda_k\pi}}{\pi\lambda_k}\right)+\tilde{\gamma}(z)\right]\tilde{\varphi}(z). \tag{35}$$

Equations (34) and (35) for z = ω are represented as follows:

$$\tilde{r}(\omega)-\tilde{\pi}(\omega)=\left[-\sum_k{}'\tilde{\Phi}_k(\omega)+\tilde{\gamma}(\omega)\right]\tilde{\varphi}(\omega) \tag{36}$$

$$\tilde{r}(\omega)+\tilde{\pi}(\omega)=\left[\sum_k{}''\tilde{\Phi}_k(\omega)+\tilde{\gamma}(\omega)\right]\tilde{\varphi}(\omega). \tag{37}$$

To determine the inverse Fourier transformation for these functions let us use the convolution theorem. We have

$$\varphi(t)=\frac{1}{2\pi}\int_{-\infty}^{+\infty}\frac{\sin\left(\pi\sqrt{\omega^2+\beta^2}\right)}{\sqrt{\omega^2+\beta^2}}e^{i\omega t}d\omega=$$

$$=\begin{cases}\frac{1}{2}I_0\left[\beta\sqrt{\pi^2-t^2}\right]; & |t|<\pi,\\ 0; & |t|>\pi\end{cases} \tag{38}$$

$$\left.\begin{aligned}&\frac{1}{2\pi}\int_{-\infty}^{+\infty}\frac{e^{i\omega t}}{\omega\pm\lambda_k}d\omega=\frac{1}{2}e^{\mp i\lambda_k t}\,sign\,t\\ &sign\,t=\begin{cases}1; & t>0\\ -1; & t<0\end{cases}\end{aligned}\right\} \tag{39}$$

And finally

$$r(t)-\pi(t)=-\frac{1}{2\pi}\sum_k{}'\int_{-\pi}^{+\pi}\left[\frac{\beta}{\lambda_k}\sin\lambda_k(\pi+t-\tau)+\cos\lambda_k(\pi+t-\tau)\right]\times$$

$$\times\,sign\,(t-\tau)I_0\left(\beta\sqrt{\pi^2-\tau^2}\right)d\tau+\frac{1}{2}\int_{-\pi}^{+\pi}\gamma(t-\tau)I_0\left(\beta\sqrt{\pi^2-\tau^2}\right)d\tau, \tag{40}$$

$$r(t)+\mathfrak{r}(t)=-\frac{1}{2\pi}\sum_{k}{}'' \int_{-\pi}^{+\pi}\left[\frac{\beta}{\lambda_k}\sin\lambda_k(\pi+t-\tau)+\cos\lambda_k(\pi+t-\tau)\right]\times$$

$$\times\, sign\,(t-\tau) I_0\left(\beta\sqrt{\pi^2-\tau^2}\right)d\tau+\frac{1}{2}\int_{-\pi}^{+\pi}\gamma(t-\tau)I_0\left(\beta\sqrt{\pi^2-\tau^2}\right)d\tau . \tag{41}$$

Adding and substracting (40) and (41) for the control functions we obtain:

$$r(t)=-\frac{1}{2\pi}\sum_{k=1}^{\infty}\int_{-\pi}^{+\pi}\left[\frac{\beta}{\lambda_k}\sin\lambda_k(\pi+t-\tau)+\cos\lambda_k(\pi+t-\tau)\right]\times$$

$$\times\, sign\,(t-\tau) I_0\left(\beta\sqrt{\pi^2-\tau^2}\right)d\tau+\frac{1}{2}\int_{-\pi}^{+\pi}\gamma(t-\tau)I_0\left(\beta\sqrt{\pi^2-\tau^2}\right)d\tau , \tag{42}$$

$$r(t)=\frac{1}{4\pi}\sum_{k}{}' \int_{-\pi}^{+\pi}\left[\frac{\beta}{\lambda_k}\sin\lambda_k(\pi+t-\tau)+\cos\lambda_k(\pi+t-\tau)\right]\times$$

$$\times\, sign\,(t-\tau) I_0\left(\beta\sqrt{\pi^2-\tau^2}\right)d\tau -$$

$$-\frac{1}{4\pi}\sum_{k}{}'' \int_{-\pi}^{+\pi}\left[\frac{\beta}{\lambda_k}\sin\lambda_k(\pi+t-\tau)+\cos\lambda_k(\pi+t-\tau)\right]\times$$

$$\times\, sign\,(t-\tau) I_0\left(\beta\sqrt{\pi^2-\tau^2}\right)d\tau . \tag{43}$$

The transfer to the initial control functions occurs for the dependencies:

$$\left.\begin{aligned} V(t) &= r(t)e^{-\beta(t+\pi)} \\ M(t) &= \mathfrak{r}(t)e^{-\beta(t+\pi)} \end{aligned}\right\}, \tag{44}$$

$$\left.\begin{aligned} U(t) &= V(t)\Delta p''-\Delta p' = r(t)\Delta p'' e^{-\beta(t+\pi)}-\Delta p' \\ F(t) &= M(t)\Delta p''-\Delta p' = \mathfrak{r}(t)\Delta p'' e^{-\beta(t+\pi)}-\Delta p' \end{aligned}\right\}. \tag{45}$$

It is quite natural that for physically minimum time the control is not the only one. Arbitrariness in the choice of the function $\widetilde{\varphi}(z)$ and the integral function $\widetilde{\gamma}(z)$ allows to impose some restrictions on the control. Therefore at passive control (a control system without feedback) the problem is practially reduced not to the complete removal of oscillations, but to the reduction of their amplitude up to the most optimal value.

Georgian Technical University

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ELECTROTECHNICS

A.Zerekidze, T.Natenadze

To the Applied Programmes Packet of DC Traction Machines Automation Design

Presented by Corr. Member of the Academy I.Jebashvili, July 4, 1996

ABSTRACT. **We report on the necessity of d.c traction electric machines automation design complex approach. The work suggests a packet of applied programmes and determines the application of each programme**.

It is known that design is an intermediate link between science and production. The application of different constructional design , change of materials and design methods significantly stipulates technical level and production quality.

The search and analysis of construction - technological solutions satisfying different demands to the designed products require complex account of all the factors determining the quality of the design.

EC provides with the possibility of saying "no" to the suggestion that all dimensions of designed electric machines and all characteristics of applied materials are always single-valued.

All the products properties and characteristics should be known with rather high accuracy only before designing experimental spesimens.

The traditional division of separate electric machines design and production doesn't satisfy up-to-date demands. Even the most perfect design fails to give us the desired results unless the demands of technology and possibilities of production have not been taken into consideration.

Should the influence of technology and quality of materials on electronic machines output parameters be envisaged, then its high quality can be guaranteed, of course on the definite level of technological culture and conformity of materials with the demands of standards.

The creation of safe machine with optimal parameters output characteristics of which are in accordance with the demands of exploitation put forward the necessesity of solving various problems of design. Thus we have to create complex system automatic design taking into account technological processes and organization of production of design.

D.c. traction machines as any other kinds of machines require programme packet creation for their automation calculations.

Such packet of programmes should at least imply the following applied programmes:

1. Parameter optimization.
2. Calculation of electromagnetic characteristics by one of the methods.
3. Calculation of possible characteristics deflections.
4. Determination of non-sparking commutation zones.

The first programme is to solve the question of chosing basic dimensions and sections of magnetic conductor and winding. With the help of the second programme which determines magnetic field distribution along all parts of the machine, electromagnetic characteristics of traction machines set with high acuracy.

If first two programmes they operate with nominal meanings of dimensions and averaged properties of magnetic and electronic conduction materials, the possible deflactions of output characteristics are defined in the third programme.

The fourth programme belongs checking calculations, when the designed machine is checked by thermal parameters.

If the second programme gives a clear picture of potential conditions on the collector, then commutational possibilities of traction machines are determined with the help of the fifth programme. Commutation qualities of d.c. machines are estimated best of all by determining their non - sparking zones.

The problem of electric machine optimal design is multicriterion, as a lot of technical and economic demands should be taken into consideration. Not only multicriterion determines the difficulties of optimal design but also unsufficient study of limited permissible quantities and interrelations of many variable parameters being in complex and contradictory dependence. That's why, for instance, in cybernetics the method of analysis and synthesis of traction electric machines [1], the way of mathematical formalization and variable agregation are chosen.

In order to use the method of mathematical formalization and generation aiming build as multicriterion corrections [2] mentioned above time on the problem deformalization and further detail design will be needed.

Optimization algorithms worked out the base of traditional methods of calculations were used, for example [3]. Some reasons within limits of permissible quantities and interrelations for such algorithm are also given.

Chosing the method in the majority of cases guided by the volume of the programme complexity of mathematical expressions and desirable accuracy of the results.

To rise the accuracy of determination characteristics for the chosen optimal variant traction, it is necessary to make calculation on strict theoretical base. The use of the electromagnetic field theory for electromagnetic calculations allows to account the satiration of separate parts of machine magnetic curcuit, their dependence and influence on each other with more accuracy. It also makes possible to arrange electric machine characteristics precisely, as they are integral characteristics of the field. This problem is solved by numeral methods in computer.

Electric machines refer to the class of products in productions of which statistic variation and probability process play essential role.

All types of electric machines including d.c. traction machines after output have been experimentally registered under test stand different from calculated level output characteristics. Moreover output characteristics of scattering meanings can be considerable.

Non homogeneity of output traction machines on technical specifications, parameters, quality indeces appear due to a number of reasons. Firstly, while producing any electrotechnical product oscillations of technological process regimes are ineritable. Secondly, in conditions of mass or serial production the oscillations of quality of rair material, half - finished products and completed products are also ineretable.

Motor output parameters (such as traction, stating moment, idle current, coefficient of performance and so on) depend on accidental meanings of all geometric dimensions and characteristics of used materials within the limits of their real scatterings, defined by fields of technological admit combination of these parameters for each specimen.

In conclusion, the results presented here indicate that calculation programmes breaking point deflactions characteristics must exist.

Georgian Technical University

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AUTOMATIC CONTROL AND
COMPUTER ENGINEERING

Z.Babunashvili, G.Popov

Automatic System for Watching Movement of Stormclouds

Presented by Academician M.Salukvadze, June 21, 1996

ABSTRACT. **Proposed system for watching movement of stormclouds together with the other methods of meteorological researches enable us to be prepared long before the storm and increase considerably the efficiency of hailprevention system.**

Our knowledge about lightning strokes based on numerous researches enables us to explain physical mechanism, character of lightning, determine trajectories of stormclouds and foresee places of strokes on to the ground [1,2].

Present work describes the automatic system for watching movements of stormclouds. The principle of detection of magnetic waves emanated from lightning is used in the system.

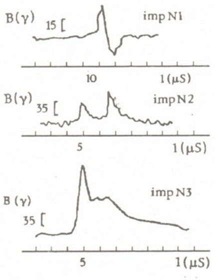

Fig.1

Fig. 1 shows oscillating currents of magnetic impulses emanated from lightning [3]. The impulses NN 1 and 2 are the results of electrical discharges of clouds. The impulse 3 corresponds to the discharge between the stormcloud character and its duration makes $20{\cdot}10^{-6}$ - $30{\cdot}10^{-6}$ sec. Frequently there were registered unipolar impulses (impulse 2) that preceed bipolar ones.

In a number of countries scientists invented similar systems reacting electrical discharges between stormclouds and the ground. For watching the front of the storm we need the information not only about strokes to the ground but also about discharges inside stormclouds and between separate clouds which are much more frequent. For this purpose we use the multichannel kilometer-length-wave detector.

The system is intended for watching the movement of the stormclouds in regions where such atmospheric phenomena as hail, storm, squale etc. are frequently observed. Such system is needful in well developed agricultural lands such as wine-growing regions in Georgia and Bulgaria.

Fig. 2 shows the scheme of the system for watching the storm front. The system consists of several multichannel detectors disposed on the protected area. Distance between detectors and their quantity depends on the precision of needed localization. For example for an area of 200-250 square km four detectors disposed in 15-20 km from each other are enough.

Each detector sends data to the dispatcher that by calculating them determines the location of storm front. This information can be used by many services such as

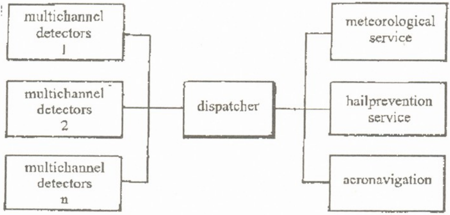

Fig.2. Scheme of the system for watching the storm front

meteorological service, dispatcher service of large energetical systems, hailprevention service, navigation, aeronavigation etc.

Each multichannel detector consists of 5 channels with different sensitivity towards east-west and north-south (Fig. 3). The system reacts to electrical discharges up to 250 kilometers away.

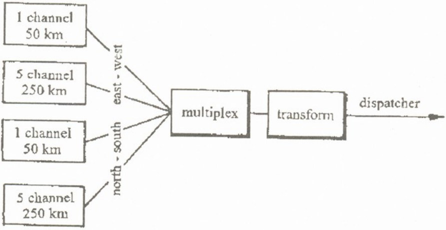

Fig.3. Scheme of the multichannel detector

The scheme of channels of the multichannel detectors is shown on the Figures 4 and 5. It presents one of the various detectors invented by our group [4].

The signal amplified by bicascade selective amplifier after straightening proceeds to the input to the comparator that determines the treshold of sensitivity of this channel. The monostable relaxation scheme excludes influence of feeble return impulses of lightning discharges. Each channel is based on an integral scheme consisting of 4 operational amplifiers with parameters similar to K 140 EL 16. The information accumulated in selection and memory schemes is transmitted to the dispatcher. The latter is communicated to the detectors so that to make them ready for receiving a new information.

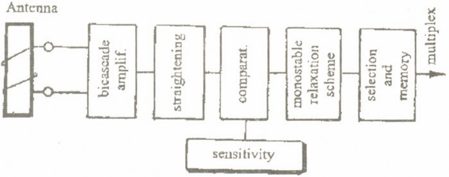

Fig.4. Scheme of the only channel

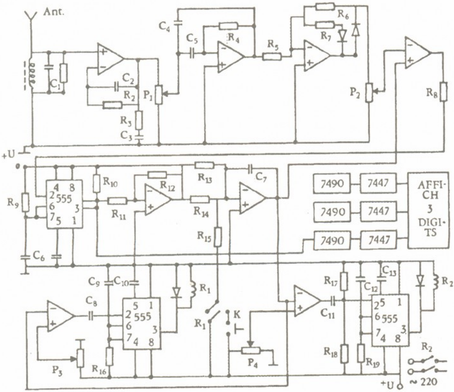

Fig.5

Precision of watching for the movement of the front of the storm depends on different factors. Among them errors caused by straight orientation of antenna, measuring technics and dissimilar group around antenna should be pointed out.

What differs proposed system of localization from other similar system is the fact that our system is able to react to electrical discharges inside stormclouds and detect all the necessary information long before the main lightning strikes the ground.

The errors point out above can be decreased by the proper choice of the places for antennae and their orientation and by excluding influence of external factors on measuring-transmitting circuits.

By increasing the amount of detectors and sensitivity of each channel we can watch movement of stormclouds for considerable distance.

In spite of its seeming immensity the system demands very simple treatment and enables us to change its separate parts without stopping its continuous work.

Institute of Electronics
Academy of Sciences of Bulgaria

Georgian Technical University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
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BIOCHEMISTRY

T.Buachidze, L.Topuria

Purification and Partial Characterization of Invertase from Fungal Fusant

Presented by Academician G.Kvesitadze, July 21, 1996

ABSTRACT. **Invertase from fungal fusant was purified by ammonium sulfate fractionation, DEAE-cellulose chromatography and Sephadex G-100 gel filtration. The purified enzyme appeared homogeneous on disc electrophoresis. Its molecular weight was estimated to be 76000 by sodium dodecyl sulfate-polyacrylamide gel electrophoresis and 75000 - by gel filtration. The carbohydrate content of the enzyme equals to 25%. The optimum activity of enzyme was observed at about 50°C and it was stable up to 55°C. Although the enzyme was stable between the pH 4.0 and 5.5 the optimum pH for its activity was about 4.6.**

Invertase (β-fructofuranosidase, EC 3.2.1.26) is widely studied in yeasts and micromycetes [1-4], and comparatively less studied in bacteria [5,6]. As far as we know there have been no reports on the purification of invertase from fungal fusant until now. This paper deals with the purification and some properties of invertase from fungal fusant.

Materials and Methods

Materials. The following materials were used in our work: DEAE-cellulose, DE-52 ("Whatman" U.K.), Sephadex G-100 ("Pharmacia Fine Chemicals" Sweden), Chemicals for electrophoresis ("Reanal" Hungary). Other chemicals used were of analytical grade and obtained from "Reakhim" firm (Russia).

Cell extract. Invertase was extracted from fusant obtained by means of fusion protoplasts from micromycetes - *Aspergillus niger* and *Allesheria terrestris* [7]. The disintegration of fusant biomass was carried out by means of plasmolyze, utilizing toluene according to the modified method of Yun [8]. The mixture was centrifuged to remove cell debris. The resultant solution was referred as the cell extract. This extract was used as crude enzyme solution.

Enzyme assay. The activity of invertase was measured by following release of reducing sugar from sucrose. The reaction mixture consisted of 0.5 ml of 6% sucrose solution, 0.4 ml of 0.05 M acetate buffer solution (pH 4.6) and 0.1 ml of diluted enzyme solution. The reaction was carried out at 50°C for 15 min. The amount of reducing sugar released was determined by the method of Šomogyi [9]. One unit of enzyme activity was defined as the amount of enzyme capable of liberating 1 μM equivalent of glucose in 1 min. Specific activity is expressed as units per mg of protein.

Protein measurement. Protein concentration was measured by the method of Lowry et al. [10] with bovine serum albumin as the standard protein. The protein in eluates column chromatography was monitored by following absorbance at 280 nm.

Purification procedure

Stage 1: Ammonium sulfate. To the cell extract, solid ammonium sulfate was added to give 45% saturation and the pH was adjusted to 4.9. After 3h the resulting precipitate was centrifuged and additional ammonium sulfate was added to the resultant supernatant to give 70% saturation. After 6 h the resultant precipitate was collected, dissolved in a small volume of buffer and ultrafiltrated ("Amicon", PM 30 membrane, USA).

Stage 2: DEAE-cellulose. The ultrafiltrated and concentrated enzyme solution was applied to a column (2.0× 29 cm) of DEAE-cellulose equilibrated with the 0.05 M acetic buffer pH 4.6. The elution was carried out with a linear gradient of NaCl (0 to 0.3 M) in 0.5 L of the same buffer. The active fractions were pooled and concentrated by ultrafiltration.

Stage 3: Sephadex G-100. The concentrated enzyme solution was put onto a Sephadex G-100 column (2.0× 65) equilibrated with 0.05M acetic buffer, pH 4.6 and eluted with the same buffer. The active fractions were combined and concentrated by ultrafiltration.

Gel electrophoresis. Polyacrylamide disc gel electrophoresis (PAGE) was done in a 12.5% acrylamide gel with 0.05 M Tris - 0.38 M glycine buffer solution at pH 8.3 [11].

Molecular weight determination. The molecular weight of the purified enzyme was estimated by sodium dodecyl sulfate gel electrophoresis [12] and Sephadex G-100 gel chromatography [13].

Estimation of carbohydrate. The carbohydrate content of a purified invertase was determined by the phenol-sulfuric acid method of Dubious et al. [14] using glucose as standard.

Results and Discussion

We have chosen the traditional methods of protein purification, namely, ammonium sulfate fractionation, ion-exchange chromatography and gel filtration. After each stage the method of ultrafiltration was used to concentrate protein and attain the pH in the solution.

We avoided the widely used method of organic solvents (ethanol, acetone, isobutanol) because enzymatic activity is partially lost. That's why we prefered the method of precipitation of protein by ammonium sulfate.

During the DEAE-cellulose column chromatography stage, invertase was recovered in one peak used as the sample for further purification. (Fig.1) After the second gel chromatography with Sephadex G-100, the specific activity of the

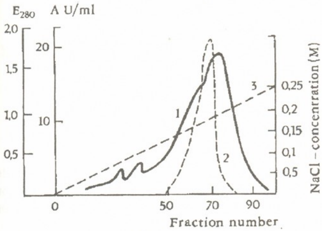


Fig 1. Eluation pattern of invertase from DEAE - cellulose column:
1 – Absorbance at 280 nm.
2 – Invertase activity.
3 – NaCl concentration.

enzyme was about 62-fold that of the crude enzyme extract.

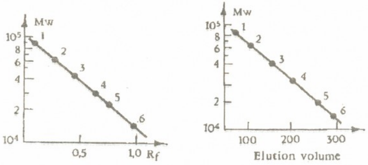
The results of the enzyme purification are summarized in the Table.

Table

Summary of purification of invertase from fusant

Purification stage	protein (mg)	Total activity (unit)	Specific activity (u/mg)	Degree of purification	yield (%)
Cell extract	875.0	4200	4.8	1.0	100
Ultrafiltration	634.0	3950	6.24	1.3	94
$(NH_4)_2SO_4$ fractionation	82.4	2925	35.5	7.4	70
DEAE-cellulose column	7.4	2014	272.2	56.7	48
Gel-filtration G-100	4.2	1250	297.6	62.0	30

The obtained preparation gave a single band on polyacrylamide disc gel electrophoresis (photograph not shown).The molecular weight of the enzyme was estimated to be 75000 by gel chromatography and 76000 by sodium dodecyl sulfate polyacrylamide slab gel electrophoresis (Fig.2). According to literary data, the molecular weight of invertase from *Aspergillus ficuum* invertase equals 84000 [15].



b a

Fig.2. Estimation of molecular weight of invertase by gel filtration on Sephadex G-100 (a) and SDS-polyacrylamide gel electrophoresis (b). The molecular weight of standard proteins are: 1 – Phosphorylase b -94000, 2 – Bovine serum albumin -68000, 3 – Ovalbumin -43000, 4 – DNA ase -31000, 5 – Mioglobin -17000, 6 – Lizocim -14000.

It's generally known, that invertases produced by almost each type of micromycetes are glycoproteins and the carbohydrate content in their molecules is 10-50%. In our case invertase molecule contained 25% of carbohydrate, while invertase from *Aspergillus awamori* contains about 50% of carbohydrates [16].

The effects of pH and temperature on invertase activity were examined. The enzyme activity was measured at various pHs (from 3.5 to 5.5) and temperatures (from 45°C to 70°C). The highest activity was observed between pH 4.0-5.0 and the temperature optimum for enzyme activity was about 50°C.

The enzyme was stable over the pH range 3.5-5.5 and up to 55°C. In addition invertase from fusant was stable at 60°C for 80 min and the remaining activity levels after 30 min at 65° was about 50%.

Thus, invertase isolated and purified from a fungal fusant has molecular weight 76000, carbohydrate content in molecule is 25%, the enzyme is stable for 80 min at 60°C, pH-stability is 3.5-5.5 region.

Georgian Technical University

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Erratum:

"Selection of Optimum Nutrient Medium for β-Galactosidaze Biosynthesis by Bacterium Thermoanaerobacter Sp.2905"

Bull. Georg. Acad. Sci., v. 153, N 3

M.Tsereteli, M.Baramidze, A.Tsereteli

The Institute of Plant Biochemistry of the Georgian Academy of Sciences, Tbilisi, 380059

1. β instead of **B**

2. On page 285, the ninth paragraph contains errors that were introduced during the production of the article. The correct paragraph is reproduced here:

"Protein concentration in culture liquid and in adjustable cell extract was determined by Lowry et al [5]. β-galactosidase activity in culture liquid as well as in cell extract was determined by Kuby and Lardy [6]. The amount of enzyme that hydrolyses 1μ mole ortho-nitrophenil-β-D-galactopyranoside at 70°C, pH-7.0 is adopted as a unit of β-galactosidase activity".

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
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BIOCHEMISTRY

R.Akhalkatsi, T.Bolotashvili, R.Solomonia, Corr. Member of the Academy N.Aleksidze

Identification of Concanavalin A Binding Proteins of Rat Brain Cellular Nuclei By Gel Electrophoresis and Blotting

Presented March 5, 1996

ABSTRACT. **The existence of concanavalin A-binding proteins has been discovered in rat brain cellular nuclei PBS soluble (180, 150, 125 and 63 kD) and triton X-100 extractable fractions (180, 175, 150, 125, 120, 107, 95, 70, 63, 57 kD) by the method of gel electrophoresis and blotting. A hypothesis is presented, that rat brain nuclear membrane pores modulator is concanavalin A lectin like protein.**

The existence of concanavalin A (Cen A)-binding glycoproteins in the rat brain cell nuclei outer membrane has been supposed [1,2]. This conclusion has been founded on the results of rat brain cell nuclei agglutination by concanavalin A [2]. Our report deals with study of identification and separation of PBS-soluble Con A-binding proteins of the rat brain cell nuclei by gel electrophoresis and blotting.

White rats of both sexes weighing 100-120 g were used as experimental objects. Rat brain cell nuclei were isolated by the method of Chauveau [3]. The purity of nuclear fractions was controlled by microscope. The nuclear suspension was centrifuged and to the nuclear pellet 20 mM potassium-phosphate buffer (pH 5.0) + 0.9% NaCl (PBS) with 5 mM phenylmethylsulphonyl fluoride in the ratio of 1 to 3 was added and homogenized (±4°C). After centrifugation (16000 g/20 min) the PBS soluble protein fraction (supernatant) was dialyzed against PBS (pH 7.4). The pellet, a crude membrane fraction was washed by PBS (pH 5.0) in order to remove fully PBS soluble proteins. Then 0.1% triton X-100 solution, prepared on 40 mM PBS (pH 7.4) was added, homogenized and the suspension was left for extraction (30 min ±4°C). After centrifugation at 6000 g/20 min the supernatant, the membrane protein fraction, was stored in vials at -70°C until use.

To identify the rat brain cell nuclear glycoconjugates, the concanavalin A was used as a lectin, which spesifically interacts with terminal glucose and/or mannose of glycoproteins. The membrane and PBS soluble proteins were dialized against bidistillate and dissolved in 5% SDS. The amount of the protein was determined by the method of Peterson [4]. The electrophoresis was carried out in the 5-12% gradient of polyacrylamide linear gel as described by Laemli [5]. In all experiments 100 μg of proteins were used. The electrophoretic transfer of proteins from the polyacrylamide gel to nitrocellulose filters (the size of pores equals 0.22 microns) was carried out according to Burnette [6]. The filters were incubated in the 10 % solution of bovine serum albumin (BSA) for 60 min. Then in a 50 mM Na-phosphate buffer (pH 7.5), containing ^{125}I -labelled Con A (100 μg/ml, 1×10^6 cpm/min), 200 mM NaCl, 1% BSA, for 2 hours at ±4°C. The filters were washed 5 times, dried and autoradiographed. The X-ray film PM-B was used for autoradiography. Parallel the specificify of the developed strips was controlled by the filter, incubated in the buffer with the same amount of proteins, containing the iodinized lectin and α-methyl-D-mannoside and α-methyl-D-glucoside in the final concentration of 0.2 M, as haptens, competitively blocking the Con A binding centers. Filters were counted in a Packard gamma counter and the specific binding defined as binding in the absence minus binding in the presence of excess (0.2 M) unlabelled haptens.

The calibrating curve was constructed by the set of standard proteins (Sigma). The standart proteins were also subjected to blotting and the electrophoregrams were coloured by Coomassie blue R-250 [6]. All other chemicals used in this study were of the highest purity available.

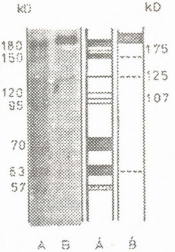

Fig. 1. The radioautograms of PBS-soluble (B) and triton X-100 extractable (A) concanavalin A binding protein and their scheme (B^1, A^1).

As a relust of the carried out experiments Con A-binding 4 glycoproteins (180,150, 125 and 63 kD) were revealed in the rat brain cell nuclei PBS soluble protein fraction (Fig. 1, B, B^1). The results of blotting experiments as a photo of autoradiography (A, B) and their schemes (A^1 , B^1) are represented in Fig. 1. One can see that among those only 180 kD glycoprotein had the highest concentration, the rest of the glycoprotein were revealed as tracks on radioautograms.

The rat brain cell nuclei membrane proteins, extracted by triton X - 100 , unlike the PBS soluble protein fractions, revealed a great capacity in ^{125}I-concanavalin A binding (Fig. 1 A, A^1). The molecular weights of Con A - positive glycoproteins varied from 57 to 180 kD, their common amount equals to 10. Three proteins with the molecular weight of 180, 70 and 63 kD are more clearly seen on the autoradiogram, seven of them belong to minor glycoconjugates (175,150, 125 107, 95 and 57 kD). It must be mentioned that all the four PBS soluble glycoconjugates were discovered in the membrane protein fraction extracted by triton X - 100. According to the literature Con A-binding glycoprotein with the molecular weight of 180 kD is discovered in the rat nuclei of liver [7]. It was proved that 180 kD glycoprotein is nuclear membrane pores protein, whose carbohydrate part is oriented inside cisternal space.

Con A -positive glycoprotein with the molecular weight of 63 kD was identified in the nuclear membrane pores by Davis and Blobel [8]., and later in the rat liver nuclei (63 - 65 kD) [9]. It was established that this 63 -65 kD protein is also a part of the nuclear membrane pores protein and plays an essential role in active transport mechanisms particularly the blocking of the nuclear pores 63 -65 kD glycoprotein with wheat germ agglutinin (WGA) inhibits the nuclear protein transport [9]. It is necessery to emphasize that the rats; intact liver nuclei are stained with FITC-WGA conjugates, while the damaged nuclei - with FITC-Con-A. These results suggest that the WGA-binding GlcNac glycoconjugate residues are localized on the external part of the nucleus membrane. This gives us possibility to employ a lectin column chromatography and preparative electrophoresis for the separation of rat brain cell nuclei membrane 63 kD protein to get further support for defining their functional role in the brain cell nuclei membrane activity. Fortunately our results represented in Fig. 1, assure us of perspectivity of the current research. In contrast with these findings, neither membrane bound, nor PBS soluble protein fractions of rat brain cell nuclei were able to bind WGA [1, 2, 10-15]. Only the slight binding trace of WGA has been discovered by us in the nuclei membrane protein fractions extracted by triton X-100. It must be noted that this ability was lost at the level of separate proteins obtained after triton X-100 extract fractionation on the column Protein PAK-300-SW [11,15].

With respect to the above-mentioned it is probably possible to conclude that Con A, but not WGA, may play a critical role in brain nuclear membrane pores modulation. The research is in progress to identify Con A like lectin as rat brain cell nuclear membrane pores modulator.

I.Javakhishvili Tbilisi State University

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BIOCHEMISTRY

I.Chkhikvishvili, A.Shalashvili, G.Papunidze, I.Targamadze, T.Meladze

Polymethoxylated Flavones and Flavanones of Citrus Flowers

Presented by Corr. Member of the Academy N.Nutsubidze, June 13, 1996

ABSTRACT. **Polymethoxylated flavones and flavanones have been investigated in Citrus flowers. It was revealed that the flowers of orange, mandarin, lemon and grapefruit contain polymethoxylated flavones: 5-OH-6, 7, 8, 3', 4' - pentamethoxyflavone, tangeretin, tetra-O-methylscutelarein, 3, 5, 6, 7, 8, 3', 4' - heptamethoxyflavone, nobiletin, sinensetin and flavanones: didimin, poncirin, hesperedin, neohesperedin, narirutin, naringin, eriocitrin and neoeriocitrin.**

One of the main peculiarities of citrus plants is high content of flavonoid compounds and their qualitative diversity [1]. At the same time it should be mentioned that chemical composition of flavonoids in citrus fruits and leaves was investigated, but in other parts the subject was less studied. Citrus flowers are of great interest as they can be used for production of nonalcoholic soft drinks [2]. It is known that citrus plants intensively blossom and a great number of flowers fell down [3]. Thus the goal of our work is to study chemical composition of polymethoxylated flavones and flavanones of some citrus flower flavonoid compounds of the varieties cultivated in Georgia. According to the existing data polymethoxylated flavones are characterised by fungistatic activity [4], but some of them (tangeretin, nobiletin) activise monooxigenase system of human liver at detoxication of carcinogenic compounds [5]. Citrus flavanones play a great role in technology of citrus processing and possess pharmacological activity [1].

Flowers of lemon "Mayer" (Citrus limonia Osbeck), lemon "Kartuli" (Citrus limon Burm), mandarin "Unshiu" (c. unshiu Marc.), orange of the variety Adgilobrivi (c. sinensis Osbeck) and grapefruit, variety Dunkan (c. paradist Mcf.) were gathered in Batumi at the experimental field of Georgian Canning industry of Research and Educational Institute.

In order to study polymethoxylated flavones 50 g of flowers of lemon "Mayer", mandarin and orange fixed by 150 ml ethanol on boiling bath, were taken. Ethanol was evaporated after filtration in a rotating evaporator. Polymethaxylated flavon extraction was carried out from resulted water solution through separating funnel 5 times with 20 - 20 ml of benzol. After ethanol extraction the remained flowers were dried in absorbing chamber. The extraction of polymethoxylated flavones from the residue was conducted by benzene on a boiling bath 3 times (150 - 150 ml). Benzene extracts were collected and evaporated to dry redisue that was dissolved in the system: hexan-isopropanol (70 : 30). Qualitative composition of polymethoxylated flavones was stated by the method of two dimensional thin-layer chromatography ("silufol" plate, 1 direction: chloroform - methanol 99 : 1, 2 direction: hexan - isopropanol (70 : 30). As an authentic samples we used polymethoxylated flavones, from mandarin Unshiu fruit skin (6). Quantitative content of flower polymethoxylated flavones was determined by high pressure liquid chromatographer (Laboratorni pristoje, Prague, Column SGX-5 m, 3×150 mm); elution was carried in a solvent system, at a rate 400 l/min, sensitivity 0.2, pressure 180mpa, detection 280 nm. Calibration curve was constructed against nobiletine [7,8] (Fig.1 a, b).

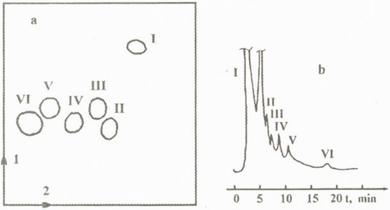

Fig.1. Separation of polymethoxylated flavones by thin layer (a) and high pressure liquid (b) chromatography methods in orange flowers.
I. 5-oxy-6, 7, 8, 3', 4' - pentamethoxyflavone, II. 5, 6, 7, 8, 4', - pentamethoxyflavone (tangeretin), III. 5, 6, 7, 4' - tetramethoxyflavone, IV. 3, 5, 6, 7, 8, 3', 4' - peptamethoxyflavone, V. 5, 6, 7, 8, 3', 4' - methoxyflavone (nobiletin), VI. 5, 6, 7, 3', 4' - pentamethoxyflavone (sinensetin).

According to the received data (Fig.1) it was found that the qualitative composition of polymethoxylated flavones of lemon "Mayer", mandarin "Unshiu" and orange "Adgilobrivi"flowers was analogous and they contained: 5-oxy-6, 7, 8, 3', 4' - pentamethooxyflavone [1], 5, 6, 7, 8, 4' - pentamethooxyflavone (tangeretin, II), 5, 6, 7, 4' - tetramethoxyflavone (tetra - O - methylscutelatein III), 3, 5, 6, 7, 8, 3', 4' - heptamethoxyflavone (IV), 5, 6, 7, 8, 3, 4 - hexamethoxyflavone (nobilitin V) and 5, 6, 7, 8, 3', 4' - pentamethoxyflavone (sinensetin VI). Proceeding from quantitative composition of polymethoxylated flavones (Tabl. 1) it should be stated that the flowers of orange "Adgilobrivi" contain twice more of these compounds, than lemon "Mayer" and mandarin "Unshiu". From separate polymethaxylated flavones in all these three plant flowers predominates 5-oxy-6, 7, 8, 3', 4' - pentamethoxyflavone, then comes tangeretin and 3, 5, 6, 7, 8, 3', 4' - hentamethoxyflavone and the rest three compounds are presented in comperatively less quantities.

To study flavanones from flowers of lemon "Mayer" lemon "Kartuli", mandarin, orange and grapefruit the analysed flower samples were fixed at 60°C by drying.

Table 1

Quantitative composition of polymethoxylated flavones in citrus flowers (mkg/g of fresh weight)

№	Polymethoxylated flavones	Orange "Adgilobrivi"	Mandarin "Unshiu"	Lemon "Mayer"
1.	5-oxy-6, 7, 8, 3', 4' - pentamethoxyflavone	299	185	190
2.	5, 6, 7, 8, 4' - pentamethoxyflavone	66	33	51
3.	5, 6, 7, 4' - tetramethoxyflavone	29	3	2
4.	3, 5, 6, 7, 8, 3', 4' - heptamethoxyflavone	45	16	4
5.	5, 6, 7, 8, 3', 4' - hexamethoxyflavone	32	14	17
6.	5, 6, 7, 3', 4' - pentamethoxyflavone	18	2	2

Flavanone extraction was carried out from 5 g of dissected material by 80% methanol on boiling bath. Extracts were collected, filtrated, evaporated and diluted with methanol. Qualitative composition extracted from flowers flavanones was determined by the method of high effective thin-layer chromatography; solvent system: nitromethan - methanol (5 : 2), 5 - 7 minutes. Chromatogram exposition was carried by 2% borhydride methanol solution in hydro-chloric acid medium. As flavone authentic samples were used hesperedin (Gee Lauson, Chemicals LTD, England), neohesperedin (Hoffman - La - Roche, Switserland), narirutin, isolated from fruit pulp of mandarin Unshui [10], naringin (Loba - Chemie, Austria) and eryocitrin, received from lemon "Dioscuria" fruit skin [11]. It should be stated that citrus plants contain two types of flavones: flavanonrutinosides, that have no bitter taste and flavanonneohesperidosides with rather bitter taste.

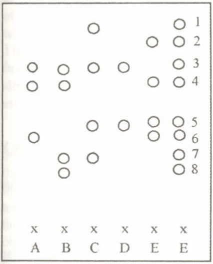

Fig.2. Flavanone separation by the method of polyamid high effective thin-layer chromatography in citrus flowers.
A - lemon "Mayer", B - lemon "Kartuli", C - mandarin "Unshiu", D - orange "Adgilobrivi", E - grapefruit "Dunkan", F - authentic samples: 1 - didimin, 2 - porcirin, 3 - hesperidin, 4 - neohesperidin, 5 - narirutin, 6 - naringin, 7 - eriocitrin, 8 - neoeriocitrin

Proceedings from the received data (Fig.2, Tabl.2) lemon "Mayer", Georgian variety of lemon, grapefruit, "Duncan" flowers contain rutinosides as well as neohesperidosides. Two neohesperidosides: neohesperidin (hesperetin - 7 - O - neohesperidoside) and naringin (naringenin - 7 - O - neohesperidoside) and one rutinoside hesperedin (hesperetin - 7 - O - rutinoside) were found in lemon "Mayer". In lemon "Kartuli" rutinosides were represented by hesperedin and eriocitrin (eriodictiol - 7 - O - rutinoside), but neohesperidosides were represented by neohesperidin and neoeriocitrin (eriodictiol - 7 - O - neohesperidoside), grapefruit flowers contain neohesperidosides: neohesperidosides:neohesperidin, naringin poncitin (isosacuranetin - 7 - O - neohesperidoside) and one rutinoside - naritutin (naringenin - 7 - O - rutinoside). In contrast to three abovementioned plants the flowers of mandarin Unshiu and orange

"Adgilobrivi" contain only flavanontrutinosides and the received soft drink has no bitter taste. Flowers of mandarine "Unshiu" contain hesperedin, narirutin, eryocitrin and didimin (isosacuranetin - 7 - O - rutinoside) and orange flowers of "Adgilobrivi" - hesperedin and narirutin. Thus citrus flowers contain biologically active polymethoxylated flavones and flavanones and can be used as the source of these compounds.

Table 2

Flavanone qualitative composition of citrus flowers

Flavanones	Lemon "Mayer"	Lemon "Kartuli"	Mandarin "Unshiu"	Orange "Adgilobrivi"	Grapefruit "Duncan"
1. Didimin			+		
2. Poncirin					+
3. Hesperidin	+	+	+	+	
4. Neohesperidin	+	+			+
5. Narirutin			+	+	
6. Naringin	+				+
7. Eriocitrin		+	+		
8. Neoeriocitrin		+			

S.Durmishidze Institute of Plant Biochemistry
Georgian Academy of Sciences

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ECOLOGY

Academician K.Nadareishvili, H.Grebenchuk

Questions of Linear Forecasting in Ecology and their Decision by Methods of Decomposition on Double Orthogonal Systems of Functions

Presented October 1, 1996

ABSTRACT. **The paper furnishes for linear forecasting and retrospective analysis of the state of technogenic fields at the decision of problems of ecodynamics.**

Shown, that, an unsoluble situation, as it were, when the forecast of parameters of technogenic fields cannot give reliable results, the opportunity of construction of the mathematical theory, allowing at certain restrictions of the character of temporary series to build the correct forecasts and to make retrospective analysis. This opportunity is associated with the methods of decomposition of finite functions on double orthogonal basis. Thus, if the basis is orthogonal simultaneously on a final piece and on the whole axis, it is possible to build estimations in past and in future with any beforehand given error.

Linear forecasting and retrospective analysis of the state of technogenic fields at the decision of problems of ecodynamics are of utmost interest, since they allow to judge the tendencies of their development and to estimate their possible consequences. Thus initial material for the construction of such estimations is only small on the extent of temporary series of the observed information, which, cannot probably allow to solve this problem with a sufficient degree of reliability. Insignificant volume of the observed information, which can be used for the decision of these problems, is due to the fact that theoretical ecology is still in a stage of formation, and also to the circumstance, that the mankind has only recently realised danger to the nature, resulting from tecnogenic activity.

Nevertheless, an unsoluble situation, as it were, when the forecast of parameters of technogenic fields cannot give reliable results, the opportunity of construction of the mathematical theory, allowing at certain restrictions of the character of temporary series to build the correct forecasts and to make retrospective analysis. This opportunity is associated with the methods of decomposition of finite functions on double orthogonal basis. Thus, if the basis is orthogonal simultaneously on a final piece and on the whole axis, it is possible to build estimations in past and in future with any beforehand given error.

Considering the observations on the parameters of technogenic fields as the measuring process, it is possible to write down for the equation of convolution

$$f(y) = \int_{-\infty}^{+\infty} F(x)\, h(y-x)dx \tag{1}$$

Where $F(x)$ is true course of parameters, $f(y)$ is the observed course of the same parameters, and $h(y-x)$ is pulsing characteristic of the measuring system.

Studying the equation (1), it is necessary to consider an opportunity of a choice of such decisions, which are correct. If one limits a class of inputs to some conditions of uniformity, it is possible to define value η^*, not dependent on individual properties of each of plausible inputs and satisfying the ratio

$$h^*(\varepsilon, h) \geq h(\varepsilon, h, F),\ \eta^* \to 0 \text{ where } \varepsilon \to 0. \qquad (2)$$

In this case it is natural to speak about uniform correctness of the equal decision (1) in a considered class of inputs. It is important to note, that it is directly due to differential properties of inputs: the more abruptly function $F(x)$ can change, i.e. the thinner the structure of an input, the greater is the error.

Proceeding from this fact, we shall hereinafter consider the forecast for a class of inputs, at which complexity changes. Convenient from both practical, and computing point of view are the classes with final number of degrees of freedom. A natural way here will be consideration of more and more complicated classes, where the complexity is understood as a subtlety of their differential structure. Such classes are built on the basis of the widely used approach to the presentation of functions as series of the orthogonal system and consideration of their partial sums. It means, that as space, on which integral operator is set, finite dimension space is chosen and then the chain of extending spaces of increasing dimension is considered. Practically, certainly, suffice it to consider finite dimensionality of spaces, whose dimensions are to be defined by desirable complexity of allowable inputs and the more the complexity we are interested in, the more is their dimension.

Let $\{Y_k(x)\}$ is some fixed system of functions, complete on an interval $(-\alpha, \alpha)$, and let $\Phi_N = \{\Phi_N(x)\}$ class of functions, represented as

$$F_N(x) = \sum_{0}^{N} \alpha_k \Psi_k(x) \qquad (3)$$

The functions $Y_k(x)$ are numbered so, that with the growth of k they become more and more complex. Therefore the highest possible value of variable $F_N^I(x)$ grows with the growth of N so, that there are the more and more complex inputs. For the considered class we shall define the maxi-mum error of deviation $\eta(\varepsilon, N, h)$, achievable for the "worst" function $F_N(x)$. This error will tend to zero together with ε for any fixed N.

Let us consider an elementary, but nevertheless practically the most interesting situation, when the input $F(x)$ is measured at a finite interval $(-\alpha,\alpha)$ with an error, characterized by mean square value, i.e. instead of $F(x)$ as a result of measurement we receive the function $F_\varepsilon(x)$, and it is known, that the error is limited by the value σ_α:

$$\int_{-\alpha}^{\alpha} \left|F_\varepsilon(x) - F(x)\right|^2 dx \leq \sigma_\alpha^2 . \qquad (4)$$

According to the general statement of the problem, mentioned above, we shall take as possible inputs Φ_N functions of a kind

$$F_N(x) = \sum_{0}^{N} \alpha_k \Psi_k(x/\alpha, c) \quad (-\alpha \leq x \leq \alpha) \qquad (5)$$

where $\Psi_k(z, c)$ are functions, satisfying the integral equation

$$\lambda_k \Psi_k(z, c) = \int_{-1}^{1} \left\{ \sin c(x-z) \Big/ \pi(x-z) \right\} \Psi_k(x, c)dx, \quad (6)$$

$\lambda_k = \lambda_k(c)$ are eigen values, and function $\Psi_k(x/\alpha, c)$ are orthogonal [1] in the interval [-1, 1], so Φ_N it is N-is n-time space, tense on $\Psi_1,...\Psi_N$. The function $\underline{F}(\omega)$ - the Fourier-image $F_N(x)$ is

$$\underline{F}_N(\omega) = \sum_{0}^{N} \beta_k \Psi_k (\omega/\Omega, c) \text{ where } \beta_k = \lambda_k \alpha_k \sqrt{2\pi\lambda_k} \quad (7)$$

As soon as a class of allowable inputs of space Φ_N is chosen, then the functions $F_\varepsilon(x)$, received as a result of measurements in view of errors, should also naturally be attributed to this class. Really, if $F_\varepsilon(x)$ does not belong to Φ_N, then "extra" components, which are not contained in representation of type (5), should be discarded, filtered, as obviously insignificantones. In the geometrical terms the whole described situation looks as follows. In a Hilbert space of functions, integrated in a square in the interval $-\alpha < x < \alpha$, is considered as finite subspace of functions, admitting representation of (7). If the function $F_N(x)$ does not lie in the subspace Φ_N, then it is necessary instead of it to consider its projection to the given subspace. Now a problem of definition of a an input $F_N(x)$, i.e., final calculation of optimum coefficients β_k, is solved by the best mean-square approximation of given function by a final sum (7). Numbers β_k are Fourer coefficients of $\underline{F}_\varepsilon(\omega)$ on system $\{\Psi_k(\omega/\Omega, c)\}$:

$$\beta_k = \int_{-\Omega}^{\Omega} \underline{F}_\varepsilon(\omega) \Psi_k(\omega/\Omega, c)d\omega = \Omega \int_{-1}^{1} \underline{F}_\varepsilon(z\Omega) \Psi_k(z, c)dz. \quad (8)$$

In other words, β_k are values of components of the projection of measured function $\underline{F}_\varepsilon(\omega)$ on a basic vector Ψ_k in subspace Φ_N. After calculation of β_k and α_k the sought meanings of an input at $|x| > \alpha$ are from (5); thus analytical continuation of functions $\Psi_k(x/\alpha, c)$ is built on the whole real axis.

We shall consider now the question of accuracy, with which meanings of an input are defined at $|x| > \alpha$ with the help of the described procedure. We shall calculate greatest possible for the given class Φ_N meaning of mean square deviation of a continued input $F_\eta(x)$ from true function $F(x)$ at a given meaning of an allowable error σ_α^2 of measurement in the interval $[-\alpha, \alpha]$. For this purpose one should find among all $F_N(x) \in F_N$ such a function of $F_N^0(x)$, for which the value

$$\sigma_\alpha^2 = \int_{-\alpha}^{+\alpha} |F_N(x)|^2 dx \text{ is given, and the value } \sigma^2 = \int_{-\infty}^{\infty} |F_N(x)|^2 dx \text{ is maximum.}$$

Using double orthogonality of system $\Psi_k(z)$ and Parseval's equation, we shall receive from (5)

$$\sigma_\alpha^2 = \alpha \sum_0^N (\alpha_k)^2 \lambda_k, \ \sigma^2 = \alpha \sum_0^N \alpha_k \tag{9}$$

As numbers λ_k decrease with the growth of k, the value of s is maximum at fixed σ_α^2, if

$$\alpha_N = \sigma_\alpha / \sqrt{\alpha}\,, \ \alpha_0 = \alpha_1 = \ldots \ \alpha_{N-1} = 0 \tag{10}$$

So that

$$F_N^0\,(x/\alpha) = (\sigma_\alpha / \sqrt{\alpha}\,)\,\Psi_k(x/\alpha,\, c) \tag{11}$$

$$\sigma^2 = \sigma_\alpha^2 / \lambda_N \tag{12}$$

It is noticable, that mean square deviation of a true input from approximated, according to the Parseval's equation, coincides with σ^2:

$$\int_{-\alpha}^{\alpha} \left| F(x) - F_\eta(x) \right|^2 dx = \int_{-\infty}^{\infty} \left| \underline{F}_\varepsilon(\omega) - \underline{F}(\omega) \right|^2 d\omega = \sigma^2 \tag{13}$$

So that the value σ^2 yields simultaneously the value of the maximum error of extrapolation. Numbers $\lambda_N(c)$, as well as the functions $\Psi_k(z,\, c)$, depend on a single parameter c. At small c numbers λ_N quickly decrease with an increase of N.

Thus, considering the fact that is far from trivial and profound: double orthogonality of basic functions in (5), it is possible rather readily and in an obvious way to calculate an error of extrapolation.

Georgian Academy of Sciences
Scientific Centre of Radiobiology
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EXPERIMENTAL MORPHOLOGY

G.Gogichadze, T.Dolidze, T.Gogichadze

Possible Reason for Tumour Cells Appearance in Hemolytic Anemias of Different Origin

Presented by Academician N.Javakhishvili, June 10, 1996

ABSTRACT. **At hemolytic anemias of different genesis side by side with hemolysis process of somatic cells fusion may take place with formation of tumour cell. Some agents may induce both fusion process and destructive effects in somatic cells by induction on plasma membranes pores of different size.**

There are hemolytic anemias of different origin. For instance, these states may be induced by some exogenic hemolytic factors: different organic and non-organic hemolytic toxins (phosphorus, phenylhydrazin, saponins, arsenicum, lead) and biotoxins (snaky venom, mushroom poisons, fungus toxins, etc.), by some medicinal preparations, radiation, by some infectious agents and by heavy burns. Besides in some cases hemolytic anemias are induced by antibodies against own tissues (autoimmune hemolytic anemia). Reason of immunization at autoimmune hemolytic anemias may be infection diseases (grippe, malaria, acute anaerobic or streptococcal sepsises, pneumonias) and some other physical and chemical factors and effects. A strong relationship exists between autoimmunity and B-cell oncogenesis. According to clinical researches malignant tumours in autoimmune hemolytic anemias in 45-47%. Observation of literature sources, as well as our experience permits to suggest that quite frequently tumour calls in autoimmune hemolytic anemias have lymphoid and macrophagal nature [1].

At the same time by the opinion of some scientists [2,3,4], some toxins, even different infectious viruses (for instance, viruses of the grippe, Rubella and human immunodeficiency virus) and carcinogenic agents may be induced both by cytolytic (destructive) effects and fusion process in somatic cells.

Such different effect of these agents on somatic cells possibly depend on size of plasma membranes pores induced by them. In case of big size of the pores irreversible changes and cytolysis take place. For instance, high dose of carcinogenic agents leads to partial increase of dikaryons and polynuclear cells, but further increase of this dose induce cellular lysis. In case of this agents tropism to immunocompetent cells may induce immunodeficite of different degree [4]. In low doses of carcinogens dikaryons are most frequently observed.

Our goal is to explain the cellular mechanism of malignant tumours development at hemolytic anemias. We have carried out electron microscopical research of 25 patients blood with autoimmune hemolytic anemias. Cells were fixed in 1% glutaraldehyde and 1% osmium tetroxide and infiltrated with epoxy resin. Thin sections were stained with uranyl acetate and lead citrate and searched visually with a TESLA BS-500 electron microscope at instrumental magnification 3000-25000.

In peripheral blood of all patients with hemolytic anemias side by side with erythrocytes of normal morphology different stages of these cells destruction were observed and ghosts of erythrocytes and membrane fragments of these cells were revealed. In 16 cases we have found bi- and multinuclear cellular formations which often contained at least one nucleus of lymphocytes and macrophages. Such calls are named heterokaryons. More often cellular structures with homogeneous nuclei were seen (homokaryons) (Fig.1). This testifies possibility of fusion of above-mentioned immunocompetent cells with other ones and each other as well. Due to karyogamic theory of carcinogenesis fusion of somatic cells in certain conditions malignant tumours may be developed.

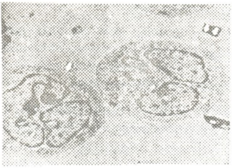

Fig.1. Peripheral blood of patient with autoimmune hemolytic anemia. Heterokaryon (→) and homokaryon (→). X 300

As leucocytes cellular membranes rigidity is more higher, it is possible that at destruction of erythrocytes (or immature cells of this line) by some agents (carcinogens, some toxins, infectious viruses, antibiotics, etc.), in leucocytes plasma membranes and pores of definite size are damaged promoting process of fusion of somatic cells. Perforations of more big size induce considerable destroy of cells plasma membranes and cytolyse (destruction) together with the perishing of this cells is followed.

Thus, at hemolytic anemias of different genesis side by side with hemolysis, process of somatic cells fusion with formation of tumour cells [5] may take place. For instance, after biting of snake with hemolytic action of venom(venom of *Vipera lebetina* and *Vipera russellii*) together with massive destruction of erythrocytes, fusion process of other cellular types with more rigid plasma membranes may be induced (for instance, leucocytes of different maturity, and probable cells of other type), with possible development of true malignant cell. Approximately such action may expect from fungus toxins (*Aspergillus flavus, Penicillum islandicum* and *Aspergillus ochraceus*) and so on. For instance, Aspergillus flavus toxin, aflatoxin induce malignant tumours of liver together with heavy toxic effect.

At autoimmune hemolytic anemias there are two possibilities of arising of malignant process: 1. After destructive action of autoimmune antibodies and immunocompetent cells on erythrocytes. It is also possible their fusogenic influence on this and other, cells, with possible development of tumour cell; 2. Some immunosupressive drugs used at autoimmune states, considerably increased risk of malignant tumours arising in particular, non-Hodgkin's lymphoma(B-cell malignancy), in resemblance of cancer of other kind. Besides, immunosupressive drugs such as cyclosporin A, chlorambucil and above all cyclophosphamide strongly increase occurence of chromosomal abnormalities in patients with connective tissue diseases. It is possible, that these substances together with cytolysis of immunocompetent cells, may induce process of fusion with formation of tumour cell.

Supposing that lymphocytes and macrophages are phenotypically dominant cells their fusion with each other and with another somatic cells may lead to tumour formation with lymphoid and macrophagal nature.

Consequently, fusion of immunocompetent cells with each other or with another ones may be regarded as possible cellular mechanism of malignant tumours formation in hemolytic anemias of different origin.

Research Center of Traumatology and Orthopedy

Research Institute of Hematology and Blood Transfusion

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EXPERIMENTAL MEDICINE

N.Okujava, Corr.Member of Academy V.Bakhutashvili, B.Korsantia

The Phagocytosis System in Patients with Destructive Pulmonary Tuberculosis and the Effectiveness of Plaferon in Stimulating of Phagocytic Activity

Presented July 19,1996

ABSTRACT. **In patients with destructive pulmonary tuberculosis the phagocytosis system was studied.Significant depression of activity of phagocytosis system in those patients was determined. The most pronounced shifts of phagocytic indices were noted in patients with fibro-cavitary tuberculosis. The rise of absorption and intracellular digestion capacity of neutrophils correlated with the severity of tuberculous process. The absence of dynamics of phagocytic indices was a significant practical value and pointed to the necessity of the use of immunomodulators along with antituberculous chemotherapy.**

Plaferon caused the positive dynamics of phagocytic indices.

Since the observation of Metchnikoff in 1891, it has been known that phagocytic cellular response is essential for competent host bacterial defence. The function of the phagocytic system depends not only on adequate number of cells and humoral factors but also on normal function of the phagocytic cells [1-7]. But till now many aspects of phagocytosis of this effective phase of struggle against M. tuberculosis are not well known.

The goal of our investigation was the to study of phagocytosis system in patients with various forms of destructive pulmonary tuberculosis and evaluation of the stimulating effect of immunomodulator plaferon on phagocytic activity.

264 patients with various forms of destructive pulmonary tuberculosis were studied. The patients were divided into two groups: 1st -130 patients receiving plaferon along with etiotropic therapy. Both groups were similar according to the age, sex, structure of clinical forms, incidences of newly detected patients and those who had been treated earlier.

Three main indices of phagocytosis system were studied: phagocyte count (phc.), phagocytic index (ph.ind.) and intracellular digestion of phagocytized bacteria (ph.dig.). Phagocytic activity of neutrophils in the peripheral blood of the patients according to Kost E.A. and Stepko M.I. [8] was determined. After the contact with staphylococcic standard cultures 100 neutrophils with engulfed mycobacteria in blood smears were counted (phagocyte count). The engulfed mycobacteria were counted and the average number of them per one neutrouphil was determined (phagocytic index). Neutrophil capacity for intracellular digestion of micobacteria tuberculosis was also studied.

The patients were examined in dynamics: on their admission, after two months of treatment and on their discharge.

Phagocytic indices in patients treated with tuberculostatics alone in regard to clinical form and results of treatment are presented in table 1. In *A* group were included patients with positive results of treatment and in *B* group - patients in whom efficiency of therapy was poor.

Table 1

The Phagocytosis system in patients with destructive pulmonary tuberculosis in dynamics

Clinical form of tuberculosis	number	Phagocytic indices	Admission	Treatment		Discharge	
				A	B	A	B
Fibro-cavitary	15	Ph count % Ph index Ph dig.%	60,4±5,2 " 2,1±0,1" 48,3±4,3"	- - -	66,3±5,9 3,0±0,1 51,4±4,7	- - --	68,1±5,6 3,1±0,2 53,3±4,1
Disseminated	30	Ph count % Ph index Ph dig.%	68,4±5,4 2,7±0,1" 53,3±4,7"	69,3±5,4 3,3±0,2 60,3±5,7	70,1±5,9 3,1±0,2 55,5±4,8	70,9±6,5 4,9±0,3 72,3±6,5	68,0±5,8 3,5±0,3 63,4±5,7
Infiltrative	45	Ph count % Ph index Ph dig.%	69,1±5,5 3,0±0,1" 56,4±5,5"	71,1±6,8 3,8±0,2 68,3±6,0""	68,2±6,0 3,2±0,1 55,9±4,4	70,3±7,0 5,1±0,3"" 71,4±6,7""	69,5±6,5 3,4±0,3 58,6±5,1
Focal	40	Ph count % Ph index Ph dig.%	68,8±5,7 2,9±0,1" 58,8±4,3"	70,5±6,6 4,,1±0,2"" 69,3±5,8""	67,8±5,7 3,4±0,1 59,7±4,7	70,5±6,3 5,5±0,4" 73,6±6,3"	68,2±6,6 3,4±0,3 60,7±5,9

Control group:Ph count-71,5%; Ph index-5,8; Ph dig. 75,7 %

(") The difference compared with control values is statistically significant.

("")The difference compared with initial values is statistically significant.

Table 2

Phagocytic indices in patients with destructive pulmonary tuberculosis treated with plaferon along with antituberculous therapy

Clinical form of tuberculosis	number	Phagocytic indices	Admission	Treatment		Discharge	
				A	B	A	B
Fibro-cavitary	19	Ph. count %	60,4±5,2 "	-	63,8±5,7	-	65,4±6,1
		Ph. index	2,1±0,1"	-	2,8±0,1	-	3,4±0,2
		Ph. dig.%	48,3±4,3"	-	52,1±4,9	--	55,3±4,9
Disseminated	30	Ph. count %	68,4±5,4	68,9±5,2	69,5±6,0	70,1±5,8	69,5±6,6
		Ph. index	2,7±0,1"	4,5±0,2"	3,4±0,1	5,1±0,3"	3,8±0,1
		Ph. dig.%	53,3±4,7"	65,8±5,7"	56,7±4,4	73,3±5,7"	65,1±6,9
Infiltrative	45	Ph. count %	69,1±5,5	70,4±6,1	68,7±5,4	71,1±5,8	70,2±7,2
		Ph. index	3,0±0,1"	4,5±0,2"	3,5±0,1	5,5±0,3"	3,3±0,1
		Ph. dig.%	56,4±5,5"	72,2±5,9"	58,5±5,7	74,4±5,6"	57,2±4,4
Focal	40	Ph. count %	68,8±5,7	69,9±6,3	70,1±6,2	70,9±6,2	69,4±5,7
		Ph. index	2,9±0,1	4,8±0,2"	3,3±0,1	5,9±0,4	3,1±0,2
		Ph. dig.%	58,8±4,3	75,2±6,5"	63,3±5,8	78,4±6,8	62,3±5,1

Control group: Ph.count-72,8%; Ph.index-6,8; Ph.dig.-79,7 %

Analysis of data showed that tuberculosis infection insignificantly influences the number of phagocytic neutrophils (exception has been noted only in the case of fibro-cavitary tuberculosis). Phagocytic count was reduced insignificantly (P<0,5). In contrast with that sharp reduction of phagocytic index and intracellular digestion of mycobacteria tuberculosis were revealed. Especially serious disturbances of phagocytosis were noted in patients with fibro-cavitary tuberculosis.

In cases with favouable course of the process, basicly referring to phagocytic index and digestion capacity of neutrophils, gradual activation of phagocytosis was observed. Pool of active cells remained on the same level. It is necessary to note that only in the case of focal pulmonary tuberculosis phagocytic index authentically rised - 5.5±0.4%, $P<0.05$ compared with initial indices - 2.9±0.1%. In cases with unfavourable course or inefficient treatment of the process significant dynamics of phagocytic indices has not been revealed.

Thus, on the basis of the obtained results it was possible to establish that phagocytic indices did not rapidly react on the changes of the patients clinical state. Long - term chemotherapy with the certain clinical effect contributes to the correction of phagocytosis defects. However, the correction of its indices were not marked in all cases, it was usually partial. Reduction of phagocytic activity as a matter of fact associated with the lack or absence of the clinical effect necessitates the use of the additional methods to influence the patients' organism, in particular with the help of immunomodulators. To restore the defects of phagocytosis mechanisms immunomodulator plaferon was used. The efficiency of this preparation in the treatment of tuberculous infection in combination with antituberculous chemotherapy was confirmed by our previous experimental investigations [8].

In patients with destructive pulmonary tuberculosis plaferon was used intramuscularly 3000 IUt in 2 ml of solution two times daily during 20 days. The results of investigation are presented in table 2.

Analysis of phagocytic indices demonstrated the rise of absorption and digestion capacity of active neutrophils, but authentically significant only in the patients of *A* group. Plaferon had no pronounced stimulating effect on the count of active phagocites. Percentage of the letters in all cases of different forms of destructive tuberculosis, in all periods of investigation regardless of the results of treatment and the use of plaferon remained at the same level.

In patients with focal and infiltrative pulmonary tuberculosis plaferon resulted in authentic rise of phagocytic index.

Thus, the results of our investigation showed significant depression of phagocytosis in patients with destructive pulmonary tuberculosis.

In patients with fibro-cavitary tuberculosis the most pronounced disturbances of phagocytosis system were revealed.

The rise of absorption and intracellular digestion capacity of neutrophils correlated with the severity of tuberculous process.

Plaferon aroused positive dynamics of phagocytic indices.

Institute of Medical Biotechnology
Georgian Academy of Sciences.

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N.Tsintsadze

HLA - Markers During Glaucoma and Cataract Diseases in Georgian Population

Presented by Corr. Member of the Academy N.Tatishvili, April 11, 1996

ABSTRACT. **Persons of Georgian nationality were examined: 69 patients having glaucoma disease and 73 patients with cataract.**

It was determined that the genetic markers of predisposition to the glaucoma disease are antigenes A1; A10; B12; gaplotipes A1-B35; A1-B5; A10-B35; B5-B12. Antigenes of cataract are HLA A3; B15; B40 and gaplotipes A11-B15; A2-B40.

Investigating genetics of various pathology great importance has the study of heterogeneity of this or that disease with antigeneities of HLA - system because the study of genetic-population of HLA antigeneities informs us about spreading of various HLA genes among different nationalities. Besides frequency prediction of this or that disease is possible (their sensibility being interchained with HLA-genes). In some cases HLA antigenes have the meaning of diagnosis.

It is known that wide spread of glaucoma and cataract in the whole world and their bad medical and social prognosis are main problems of medicine. The above mentioned has a great practical meaning for investigation of different risk factors, cure and prophilaxis of these pathologies [1,2].

The main part in formation of glaucoma and cataract has no doubt genetic determination, but its concrete mechanisms for today are not fully investigated (3,4,5). As for literature information about connection of HLA-system with these glaucoma pathologies are controversal and sometimes are not right [6-12].

From this respect we investigated HLA-antigenes "batary method" with the help of the standard two-stage microlimphotoxical test after Terasari and Mc.Clelland (1964) (coloured by trephan blue or airon). The typical system was made with HLA-antiserum of Leningrad Scientific Institute of Hematology and Blood Transfluse and commercial serum of German Firm "Behring". 28 antigenes of the I class were investigated.

In the result there were revealed persons with high risk of glaucoma and cataract (risk-groups) giving chance for prophilaxis of these deseases. Besides we paid attention to the existance population peculiarities association between HLA-complex and glaucoma and cataract in Georgian population as Georgian population is not dependent on the ethnic territory and is characterized by various features of spreading HLA-antigenes [13,14].

While studying antigenes spread during cataract in Georgia we consider that immunogenetic markers in Georgian population are antigenes: HLA A1, A-10, B[12]; gaplotipes A1-B5; A1-B35; A10-B5; A10-B35. Genetic-markers of glaucoma are HLA-antigenes A1, A10, B12 and gaplotipes A1-B5, A10-B35.

Tbilisi State Medical University

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
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EXPERIMENTAL MEDICINE

E.Pataraia, Ch.Baumgartner

Clinical Significance of Ictal EEG Patterns in Patients with Mesial Temporal Lobe Epilepsy

Presented by Corr. Member of the Academy V.Mosidse, August 22, 1996

ABSTRACT. **We analyzed lateralization accuracy and reliability of EEG seizure patterns in 118 seizures in 24 patients with mesial temporal lobe epilepsy as documented by hippocampal atrophy on MRI and unitemporal spikes on interictal EEG. Two blinded electroencephalographers independently determined morphology, location and time course of ictal EEG changes. A lateralization was possible in 88.4 to 92.0% of seizures. Lateralization always corresponded to the side of the interictal spike focus and of hippocampal atrophy on MRL. Lateralization was never incorrect. Lateralization at seizure onset was possible in only 30.4 to 33.9% of seizures. A later significant pattern which allowed lateralization occured in 83.8 to 93.6% of seizures with a non-lateralized onset and started 12.8 to 13.2 seconds after EEG seizure onset. Interobserver reliability for lateralization was excellent with a kappa value of 0.85. Concerning the individual patients, in 19-20 patients all seizures were lateralized, in 4 patients a mixture of lateralized and non-lateralized seizures occured and in one patient no seizure was lateralized.**

Introduction

In recent years the syndrome of mesial temporal lobe epilepsy (MTLE) has been delineated and distinguished from other forms of temporal lobe epilepsy. MTLE can easily be diagnosed on the basis of clinical history, anterior temporal spikes on scalp-EEG and the appearance of hippocampal atrophy or sclerosis on MRI scan [1,2]. Patients with MTLE frequently have medically refractory seizures and are considered excellent candidates for epilepsy surgery [1,2].

Presurgical evaluation of epilepsy patients relies on independent and converging evidence from history, clinical seizure semiology, interictal and ictal EEG, neuropsychological testing, the intracarotid amobarbital procedure (IAP) as well as structural (MRI) and functional imaging studies (SPECT and PET) [1,3,4,5]. Prolonged video-EEG monitoring has been the cornerstone of presurgical evaluation [6-8]. Although interictal EEG has proven to provide valuable lateralizing information [9,10], ictal EEG recordings are generally considered more reliable for the localization of the seizure onset zone centers [6,7,11,12]. Several studies have addressed the lateralizing and localizing value of various ictal EEG patterns [13-21]. Specifically, interobserver reliability [19,20], the clinical significance of ictal EEG changes in patients with uni- and bitemporal independent interictal spikes and the correlation with depth electrode recordings have been systematically investigated. However, these studies did not differentiate between the various subtypes of temporal lobe epilepsy and specifically did not separate out patients with MTLE.

In some recent articles, ictal EEG findings in patients with mesial temporal lobe epilepsy were reported [21-22]. However, these studies did not present a detailed analysis of ictal EEG patterns, they did not distinguish patients with uni- and bitemporal interictal discharges and finally did not assess interobserver reliability by blinded EEG analysis.

We analyzed systematically occurrence and lateralizing reliability of various EEG patterns in patients with unilateral mesial temporal lobe epilepsy, then we assessed the interobserver reliability of these EEG changes. And finally, we analyzed the ictal EEG changes for the individual patients which is of clinical relevance for the decision process in presurgical evaluation. Specifically, we investigated the clinical significance of non-lateralizing ictal EEG patterns.

Methods

Patients

We studied 24 consecutive patients (10 women, 14 men; mean age: 33,6 years, range 19 to 54 years) with medically refractory mesial temporal lobe epilepsy. Mesial temporal epilepsy was defined by typical seizure semiology, interictal spikes with a maximum over the anteromesial or midtemporal electrodes (FT 7-10, T 7-10, SP 1-2) and hippocampal atrophy or sclerosis on MRI scan.

EEG recording

All patients underwent prolonged video-EEG-monitoring for a definite localization of the epileptogenic zone for an average of 5 days (range: 4-14 days). Anticonvulsant medication was reduced or completely withdrawn to facilitate seizure occurrence [2,3] EEG was recorded from gold disk electrodes placed according to the extended International 10-20 System with additional temporal electrodes [24] and from bilaterally placed sphenoidal electrodes. EEG signals were recorded referentially against Pz, amplified, filtered with a bandpass of 0.3-70 Hz, analog-to-digital converted with a sampling rate of 256 Hz (12 bit) and stored digitally for off-line data analysis. A commercially available EEG monitoring system was used for data acquisition and analysis (Pegasus Monitoring System, EMS Company, Korneuburg, Austria) which allowed reformating of the data in any desired montage [25].

Interictal EEG analysis

The frequency and location of interictal spikes was assessed by visual analysis. At least two minutes of artifact-free EEG recording per hour were reviewed. A minimum of 100 spikes across the various stages of the sleep-wake cycle were analyzed in each individual patient . Only patients with unitemporal interictal spikes were included in the present study. According to previous studies unitemporal spikes were defined as a ratio of more than 90% of spikes occuring over the more affected temporal lobe [10].

Ictal EEG analysis

The following montages were used for ictal EEG analysis: 1.Longitudinal bipolar montage. 2.Transverse bipolar montage. 3.Referential montage where all electrodes were referenced against PZ.

Similar to previous studies [19] the following parameters were assessed for each EEG seizure pattern:

1. Time course of ictal discharge:

a) pattern at onset (PAO)--the first unequivocal ictal EEG change lasting for at least 3 seconds was defined as the pattern at onset.

b) later significant pattern (LSP)--a LSP was defined when a significant change in the morphology or location of PAO occurred.

2. Location of ictal discharge:
- temporal, left or right;
- hemispheric, left or right;
- bilateral, higher on the left or right
- bilateral, non-lateralized.

3. Lateralization of ictal EEG.

Only temporal (a) and hemispheric (b) ictal discharges were considered as sufficient for lateralization. Both bilateral patterns (c and d) were regarded as non-lateralized. The ictal EEG lateralization was classified as correct if it corresponded to the side of atrophy or sclerosis on MRI scan and to the lateralization of interictal spikes.

Interobserver reliability

All ictal EEG patterns were analyzed by two independent experienced EEGers (C.B. and E.P.). Seizures were reviewed by digital EEG (all options of reformating, digital filtering, gain adjustment etc.) in random order generated by a random generator across all patients and seizures; thus seizures from the same patient generally were not reviewed serially. Only the time of clinical onset was marked in each EEG file, whereas the EEGers were blinded for all other data (e.g. patient name, results of interictal EEG, side of atrophy on MRI). We used the kappa statistic in order to assess interobserver agreement [26].

Results

We reviewed 118 seizures in 24 patients, 6 seizures had to be excluded from further analysis because of artifacts. Therefore, 112 seizures were included in the final analysis. The number of seizures for each individual patient ranged from 2 to 11 (average: 5 seizures).

Pattern at onset (PAO)

Concerning location, the PAO was classified as regional over either temporal lobe in 36(32.1%, EEGer-1)resp. 31 seizures (27.7%, EEGer-2), as lateralized over one hemisphere in 2 (1.8%, EEGer-1) resp. 3 seizures (2.7%, EEGer-2), as non-lateralized with a maximum over one hemisphere in 9 (8.1 %, EEGer-1) resp. 11 seizures (9.8%, EEGer-2) and finally as non-lateralized in the remainig 65(58%, EEGer-1) resp. 67 seizures (59.8%, EEGer-2). Thus lateralization from the pattern of onset was possible for 38 (33.9%, EEGer-1) resp. 34 seizures (30.4%, EEGer-2). For the individual patients lateralization from the PAO was possible for all seizures in one patient (4.2%, EEGer-1) resp. two patients (8.3%, EEGer-2), for at least one seizure in 13 patients (54.2% EEGer-1 and 2) and for no seizure in 9 (37.5%, EEGer-1) resp. 10 patients (41.7%, EEGer-2).

Later significant pattern (LSP)

Per definition a later significant patters (LSP) was considered only for 74 (EEGer-1) resp. 78 (EEGer-2) seizures where the PAO was non-lateralized. A LSP was observed in 62 (83.8%, EEG-1) resp. 73(93.6%, EEGer-2) of these seizures. The LSP started at an average of 13.16±9.58 seconds (EEGer-1) resp. 12.81±7.60 seconds (EEGer-2) after EEG seizure onset (median: 10 seconds (EEGer-1) resp. 12 seconds (EEGer-2); range: 4-59 seconds (EEGer-1), 4-39 seconds (EEGer-2).

The LSP was regional temporal in 60 (96.2%, EEGer-1) resp. 64(87.7%, EEGer-2) seizures, lateralized over one hemisphere in 2(3.2%, EEGer-1) resp. 7(9.6%, EEGer-2) seizures and non-lateralized with a maximum over one hemisphere in 2 seizures (EEGer-2, 2.8%). In all 62 (EEGer-1) and 71 (EEGer-2) seizures with a lateralized LSP, this pattern was sustained lasting for >10sec.

Lateralisation from ictal EEG - correlation with interictal EEG and MRI-scan

A definite lateralization from ictal EEG was possible in 89 (88.4%, EEGer-1) resp. 93 (92.0%, EEGer-2) seizures. The side of lateralization was always correct, i.e. corresponded to the side of the interictal spike focus and of hippocampal atrophy or sclerosis on MRI scan. Conversely, in13 (11.6%, EEGer-1) resp. 9(8.0%, EEGer-2) seizures a lateralization was not possible from ictal EEG. Concerning lateralization for the individual patients we obtained the following results: In 19 (79.2%, EEGer-1) resp. 20(83.3%, EEGer-2) patients all seizures were correctly lateralized. In 4 (16.7%, EEGer-1 and EEGer-2) a mixture of lateralized and non-lateralized seizures was observed. In one patient (EEGer-1, 4.2%), no seizure could be lateralized.

Interobserver reliability

Interobserver reliability for lateralization considering the information derived from both the PAO and the LSP was excellent (kappa = 0.85). Accuracy for PAO lateralization was fair (kappa = 0.59) and for LSP - it was good (kappa = 0,69).

Conclusion

Our results indicate that in patients with unilateral mesial temporal lobe epilepsy a correct lateralization of ictal EEG patterns corresponding to the side of hippocampal atrophy is possible in most seizures and in patients with an excellent interobserver reliability, that lateralization - if defined by strict criteria - is never incorrect; finally in rare instances of patients with non-lateralized seizures reevaluation of MRI and interictal spikes for evidence of bilaterality should be performed. This is the first study a dealing with systematic analysis of ictal EEG patterns in patients with unilateral mesial temporal lobe epilepsy including assessment of interobserver reliability.

We used bippocampal atrophy and sclerosis on MRI scans and unitemporal interictal spikes for correct and incorrect lateralization of ictal EEG. Thus, in contrast to previous studies using successful epilepsy surgery as the criterion for identification of the epileptogenic zone our study cannot be used to derive any conclusions concerning postoperative outcome. However this should have no implications on the validity of our results because several reports have documented excellent prognosis in patients with unitemporal spikes and ipsilateral atrophy or sclerosis on MRI scan. Furthermore, 15 of our 24 patients (63.5%) were rendered seizure-free by a selective amygdala-hippocampectomy with a mean postsurgical follow-up of 18 months (range: 5 to 25 months).

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Tbilisi State Medical Univerisry

Universitätsklinik für Neurologie
Vienna, Austria

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საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №1, 1996
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PALAEOBIOLOGY

K.Chochieva, N.Mamatsashvili

Spore-Pollen Complexes of Chaudian Deposits Bearing the Decidua of Colchis Redwood Forest

Presented by Academecian L.Gabunia 19.03.1996

ABSTRACT. **Recent spectra of layers bearing the decidua of redwood forest of Colchida are presented. Data of palynological analysis as well as paleocarpological testify Colchida's antiquity as a refugium, its connection with tropical and subtropical regions in geological past and ultimate delayed rate of its organic world transformation through anthropogene.**

Conifers have been constantly observed in the vegetative cover of Georgia since late Triassic - Early Jurassic. Nevertheless the fossil fall off or decidua of the coniferous forests are rarely noticed except for pollen. Thus, the accumulation of small plant remains consisting of cones, cone scales, seeds, needle, leafy shoots of *Taxodiaceae* and *Cupressaceae* in compact fine sandy clays of the Chaudian basin (1.2 -0.7 m.y), is especially noteworthy [1]. The bedrock was so saturated with remains of conifers, especially with cones, needles and seeds of *Sequoia* and *Cupressus*, that it was impossible to clean any sample taken for analysis without its damage. *Sequoia* cones were usually found opened, scaled off, while *Cupressaceae* cones were closed. During the first summer expeditions, through a thin layer of transparent water of silted up brook one could still see big branches of these taxa deprived of any traces of transportation. However, any attempts to cut out or strike off a sample resulted in its crumbling. Faint trace of the branches on the rock vanished merging with the colour of the clay becoming bluish-grey in the air. There was no doubt that alocation of fossilization of coniferous forest decidua has been revealed. Thus the first and the only finding for today is coniferous forest of fossil decidua in the Caucasus. The decidua of the coniferous forest is compared to the type of the coast redwood of California [2]. According to the paleocarpological analysis, in the basin of accumulation (plant remain carrying layers) the stand of *Sequoia* and *Cupressus* trees predominated, whereas the palynological data have shown prevalence only of Sequoia (67-80%). In spite of the fact that there were plenty of cones, seeds and fragments of branches, *Cupressaceae* were represented only by single pollen grains. Though in the spectra of these deposits leafy stalks of moss were represented in every weight of the organic part of washed samples. Bryales were not found in spores at all. This might happen due to lack of unsufficient samples analysis. Time passed and palynological study of these deposits was renewed. In the literature available for us there was no information about a decidua *Taxodiaceae* existent out of their present-day habitat for such a long period of time, such a variety of family taxons at any late Pliocene and all the more in Anthropogene flora. For the sake of getting minutely data and checking the previous ones the sediments conventionally were subdivided into three layers according to the spreading roughly some 0.3 m dividing them from each other. A sample has been taken from each layer and from each sample a weight of some 40-60 gramms was taken for analysis. The received spectra were characterized by stable taxon composition for the Chaudian horizont spectra as well as for all paleo-palynocomplex of Colchida: various trees dominating, plenty of coniferous pollen, wide variety of Taxodiaceae, small amounts of pollen, while very wide combination of Angiospermae taxons. Somewhat unexpected was the absence of *Podocarpus* and *Abies* pollen which are constantly present in all pollen spectra of Cenozoic flora of Georgia and accordingly was presented in the same Khvarbeti florula [3]. Different appeared to be the data on presence in the roch pollen those of *Sequoia, Taxodium, Cryptomeria*. The first samples analyzed showed *Sequoia* somewhat from 26- to 80% of the spectra than the newly derived picture showed *Taxodium* as the leading taxon. All the data obtained witnessed that considerable consistence of *Taxodium* and *Cryptomeria* pollen are characteristic not only of old Euxine deposits of the West Georgia [4].

The results of palynological analysis appeared to be so much significant that it was decided to show not only the taxons variety composition but quantity and correlation of pollen in spectra as well.

Table

List of Taxons	Sample numb.	No 1 %	Sample numb.	No 2 %	Sample numb.	No 3 %
Tsuga diversifolia	17	3.6	7	4.0		
T.canadensis	18	3.8	11	6.2	19	7.1
T.sp.	19	4.0	8	4.5	6	2.2
Picea			2	1.1	1	0.3
Pinus	6	1.2	3	1.7	7	2.6
Taxodiaceae	37	7.8	22	12.5	22	8.2
Sequoiadendron	4	0.8				
Sequoia	21	4.4	6	3.4	18	6.7
Metosequoia	4	0.8	6	3.4	1	0.3
Taxodium	215	45.5	59	33.7	116	43.4
Cryptomeria	40	8.4	9	5.1	18	6.7
Cunninghamia	4	0.8				
Glyptostrobus	7	1.4	6	3.4	10	3.7
Cupressaceae			1	0.5		
Cupressus	7	1.4	1	0.5	2	0.7
Libocedrus	4	0.8	2	1.1	1	0.3
Juniperus	3	0.6	1	0.5	1	0.3
Salix					1	0.3
Populus	1	0.2				
Pterocarya	3	0.6	4	2.2	2	8.7
Juglans	9	1.9	2	1.1		
Engelhardtia					1	0.3
Platycarya	1	0.2	1	0.5		
Carpinus caycasica	1	0.2			2	0.7
Corylus	6	1.2			1	0.3
Betula					1	0.3
Alnus	2	0.4	2	1.1	4	1.4

(Continued)

Castanea	2	0.4	1	0.5	9	3.3
Quercus			1	0.5	4	1.4
Ulmus	5	1.0			1	0.3
Zelkova					1	0.3
Liquidambar			1	0.5		
Rosaceae					1	0.3
Phellodendron					3	1.1
Aesculus	6	1.2	1	0.5	5	1.8
Nyssa	7	1.4				
Fatsia	3	0.6	2	1.1		
Ericaceae	4	0.8	1	0.5		
Undetermined	16	3.3	15	0.5	9	3.3
Gramineae	1		4		3	
Chenopodiaceae						2
Nyphar						1
Umbelliferae	2		1			
Bifora	2					
Plantago	1					1
Compositae	3		5		1	
Artemisia	2		1		4	
Undetermined						
Sphagnum	1	0.3				
Polypodiceae	260	92.5	62		159	90.8
Onaclea	5	1.7	6		8	4.5
Pteris					1	0.5
Polypodium	8	2.8			4	2.2
Osmunda	3	1.0	1		3	1.7
Cyatheae	2	0.7				
Undetermined	2	0.7	1			

The composition of *Taxodiaceae* and *Cupressaceae*, presence of *Theaceae, Araliaceae, Ruraceae* and *Symplocaceae* were somewhat separating the Khvarbeti florula. But as soon as in the Chaudian deposits (the left bank of the river Tchakhvata and the vicinity of Japhareuli village) the presence of Sequoia and Athrotaxis cones, seeds and fruits of *Eurya, Fatsia, Phellodendron* and *Symplocos* were revealed the bordering lines of the florula have become more unstable and misty. It has appeared an integral whole, natural of the Chaudian flora of Guria- rich, woodsy, different from its contemporaries of Europeans, Japan's, North America's flora considering many "exotics" - the relicts of the ancient flora of the same territory. Echoes of some early connections, geologically deep roots were quite close to tropical and subtropical vegetations. We have noted before the "noncorrespondency" [5] of the Chaudian flora compound according to the layers bearing the leftovers. In any case they comperatively remote in time, the Khvarbeti florula, is one of the brightest, ever possible and sure witness of such a combination and it testifies antiquity of Colchida as a refugium.

L.Davitashvili Institute of Palaeobiology
Georgian Academy of Sciences

V.Bagrationi Institute of Geography

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HISTORY OF SCIENCE

L.Razmadze, Corr. Member of the Academy D.Jokhadze

Darwin's Evolutionary Theory and "Kvali"

Presented August 29, 1996

ABSTRACT. **The Georgian weakly political, scientific and literary newspaper "Kvali" (1893 - 1904) was one of the printed organs in Georgia publishing the most complete scientific articles on Darvin's evolutionary theory. The material described in this paper gives chance to the wide sections of population of Georgia to know about new achievements in different sciences as well as those of biology through newspapers.**

Charles Darwin's evolutionary theory that reliably explained existing living nature's variety-animals and plants and established the order of organic world's historical development is considered one of the biggest achievements in the intellectual life of the mankind. It had to deny previously believed ideas and instill new ones. The whole period after Darwin biology particularly the achievements of the last several decades certify the truthfulness of the principles expressed by this remarkable investigator and thinker. This is also proved by the results of investigations of life phenomenon on the molecular level [1, 2].

The above mentioned theory which was first published in 1859 drew the civilized world's attention and very soon it became the subject of discussions not only for specialist-biologists but of the whole society. Prominent Georgian publishmen tried to give chance to the bulk of the population of Georgia to get to know about fresh achievements in different sciences and biology through newspapers. They had to do this because at that period (until 1918) there were no high educational institutions in Georgia, where it would be possible to study sciences thoroughly. When we began to study the history of spreading the knowledge of different sciences and biology among them and especially the spreading of information about Darwin's evolutionary theory, we found out that one of the newspapers very active in this aspect was "Kvali"- a political, scientific and literary weakly newspaper, which had been published from 1893 to 1904. It was in this newspaper, where in 1896 (N 43) an information about the edition of a French author E. Feriera's new book "Darwinism" was published in Georgian translation by I.Pantskhava. "Kvali" considers the eddition of this book to be a very pleasant phenomenon and points out that "this book is widely known in Europe and is translated into all the languages of cultural countries. How astonishingly deep and interesting this theory is, the reader will understand from this book".

We must mark off the series of articles that was published in "Kvali" in several issues of 1897 (N N 10-15) with the same title "The development of animals' embryo" [4]. These articles belong to well known Georgian publicist I.Pantskhava. At first the author characterizes the achievements of science of that period and notes that the science developed mainly in XVIII-XIX centuries because of finding true facts "that defeats all different false believes created by ignorant and foolish mind, those false believes which are supported, defended and cared by religion". The achievements of science "...widened and recognized the whole world, nature and everything what it consists of as a true fact, develops and changes".

Everything grows, changes and develops. This fact is expressed by the author with the word "evolution". He thinks that life is connected with the proteins and marks that "in the eighties protein was received in chemical laboratory". It seems that the author points to the protein (polypeptide) syntheses experiments by the German scientist E.Fisher. The scientist thought it was possible to join 18 different aminoacids through peptide joints. Afterwards Fisher's follower E.Abdelharden made the number of aminoacids up to 19. At the same time the author tries to connect organic substances with inorganic because he believes that it's quite possible to create organic beings from inorganic substances.

In several issues of "Kvali" of 1899 (N N 42, 44, 45, 49 and 52) there are series of articles with general title: "From Copernicus till Darwin" adapted by P.Surguladze from different foreign texts [5]. These articles revise general achievements of different sciences for that period, they characterize the meaning of great natural scientific discoveries. The last two articles were completely dedicated to the description of main features of Darwin's evolutionary theory. These articles were published as a book with the same title in 1900.

Three rather large articles published by "Kvali" in 1902 are very interesting from the point of describing Darwin's theory [7]. Each article occupies half a page. We can see that the author is highly informed in this matter. The articles are signed with symbols XX.

In the first article which is preceded by Darwin's rather big portrait, the author notes that "beginning from the second half of the last century dominate two very strong theories from the scientific point of view and its practical importance. Both these theories became the stars for the most vivid and enthusiastic part of mankind. One theory concerns the society events and the second is called Darwinism (In the first theory Surguladze means Marxism. - the authors). Darwin's theory or Darwinism is a method, research mode, that is necessary for all: for nature scientist, for a historian, for a law specialist and for a moralist. None of human mental activities is left outside the influence of Darwinism". The author tells us Darwin's biography in a very exciting way. He describes Darwin's family genealogy. It is said that Darwin's grand father Erasmus Darwin wrote and edited a composition with the title "Zoonomy", where ideas about the ways of developing nature are given. The author tells us the story about Darwin and the other well known nature scientist Alfred Walles. He tells how Alfred Walles came to an idea about the law of natural selection independently from Darwin. At the end Walles himself recognized Darwin's priority in discovering the law of natural selection.

In the second article the author discusses the question of natural selection and its main attributing factor - that is fighting for existence as the matter of complex intercourses among living organisms. At the same time he gives his arguments. Particularly he marks off that the organisms are greatly inclined towards multiplying and together with this they multiply far more than the existing conditions can allow. This is the reason, why the fighting for existence arises, when a great number of creatures perish. Only the best adapted, those who have some advantage against the others in the given concrete conditions may preserve. In this article the author reminds us of Darwin's idea about the fact that struggle between the individuals of the same kind is more severe than that of the other species. It is known that for this idea Darwin was criticized not very long ago, because of official outlook in the former Soviet Union. Darwin was accused to be under the influence of English economist and theologian Malthus and this was considered Darwin's great mistake. We should say here that neither Maltus nor Darwin's theory of evolution has anything to do with the false official Soviet commentaries. Soviet commentaries always were in accordance with the points of view of representatives of the establishments.

The author of the article thinks that the principle of struggle for existence is true for the human society too, because in the society population increases much quicker than the means for surviving.

In the third article the author analyzes the problem of natural and artificial selection of animals and plants and practical agriculture which helped Darwin to discover the law of natural selection. The author of the article underlines Darwin's thesis that in artificial selection the changing is directed by a man in his own benefit. In natural selection the changing is directed by nature itself and it brings benefit to the organism itself: in struggle for existence only those features are left which are beneficial for organism. He recites Darwin himself: "A man selects in his own benefit, but nature for the benefit of organism. The author resumes: "Fixing of all useful changes, perpetual struggle, changing dying off of weak and defending of adapted- that is the way that nature follows. The way organisms are changing and developing little by little.

As already was said above, all three articles in "Kvali" are not signed or better to say are signed by XX. It seems that they belong to famous Georgian writer and publicist George Tsereteli, who usually signed his writings in this manner [8]. He was well acquainted with the problems of natural science. The first information about Darwin and his theory belongs to him [9, 10].

All said above certifies that newspaper "Kvali" is one of those Georgian periodicals which was eager to inform Georgian society about civilized worlds scientific knowledge (in our case - biology). There is no doubt that by this the soil was being fertilized for nurishing the high-level educational institution, for founding the University of Georgia.

S.Durmishidze Institute of Plant Biochemistry
Georgian Academy of Sciences

I.Javakhishvili Tbilisi State University

Gurjaani Agricultural Institute

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Corr. Member of the Academy T.Sanadze

To the 30th Anniversary of Discrete Saturation Pulsed EPR Spectroscopy

Presented August 6, 1996

ABSTRACT. **History of discovery and development of the discrete saturation pulsed spectroscopy and examples of its application are presented. Invited lecture on the XXVII Anniversary Congress AMPERE, Kazan, 1994 is offered.**

I'll tell you about pulsed spectroscopy based on burning of multihole spectra in EPR lines. This method was discovered and developed in Tbilisi State University and this year it is exactly 30 years of age. Creation of this spectroscopy, as you will see, is tightly connected with E.K.Zavoiski. Recently hole burning spectroscopy was modified by Prof. A.Schweiger who performed Fourier transform of such spectra and now it finds ever-widening application. In connection with this date I would like to present you the historical development of multihole burning spectroscopy. I think, it may be of interest.

In 1954, when EPR was only 10 years of age and we all were forty years younger, I came to the laboratory of Prof. A.M.Prokhorov in the Lebedev Institute as a post-graduate being sent for some months to study the EPR technique in order to continue experiments at low temperatures in Tbilisi. At that time a cryogenic laboratory at our University founded by Prof. E.Andronikasvili has been created. In Moscow I received a scheme of a superheterodyne EPR spectrometer worked out by A.Manenkov and A.Prokhorov and as a present of main components of my future hand-made spectrometer. I also got from them fluorite single crystals with the uranium impurity. In 1956, when my EPR spectrometer began to work at low temperatures and I obtained the first spectra of uranium in fluorite, I immediately wrote about it to Prof. A.Prokhorov. At once I received from him three sheets preprint by B.Bleany and his co-workers [2], in which this spectrum was investigated in detail. In 1955 Prof. A.Prokhorov and Prof. B.Kozirev visited Prof. B.Bleany in Oxford and agreed to send each other new works. So at that time I had to stop my work at the spectrum. In two years I returned to it to investigate relaxation processes and a fluorine hyperfine structure which was recognized complex by B.Bleany himself and remained unexplained. You can see this spectrum in main orientations of the magnetic field in relation to fluorite crystalline axes (Fig.1). In other orientations the structure is not resolved. For many years I tried to describe this

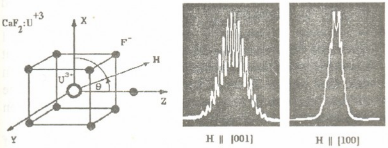

Fig.1. The magnetic center tetragonal Symmetry U^{3+} in CaF_2 according to B.Bleany model and oscillograms of the EPR lines in orientation H|| [001] and H|| [100]. Oscillograms are taken from article [5].

structure to understand what kind of splitting level took part in its formation, but without result.

Some of my colleagues dissuaded me against examining this spectrum, as they thought it was a mere waste of time, but to do so was beyond my power. This simple and at the first sight unimportant problem (explanation of a particular spectrum), became very important for me and from time to time it compelled all my attention.

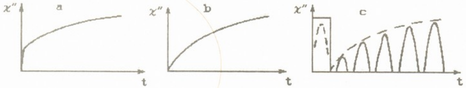

Fig.2. A typical shape of restoration of the absorption curve in EPR line at the pulse saturation: a) pulse duration is shorter than crossrelaxation time in EPR line; b) Pulse duration is longer than crossrelaxation time; c) the same at the saturation of the whole line at switched on magnetic field modulation (modulation frequency 50 Hz).

In investigation of relaxation processes in these lines by a pulse saturation technique the absorption restoration curve consisted, as usual, of two parts: a short-period and a long-period (Fig.2 a). They corresponded to cross-relaxation inside the line and to spin-lattice relaxation. It was easy to obtain a pure long-period process by lengthening the saturating pulse (Fig.2 b) or saturating the whole line at switched on magnetic field modulation. We could not obtain pure cross-relaxation process by pulse shortening and to find out the hindrance. We applied a short pulse at switched on magnetic field modulation. The picture obtained was rather impressive. The pulse was nonsynchronized with the magnetic field sweep in that moment and slowly floated along the line. At the same time the EPR line "breathed" as if it were alive.

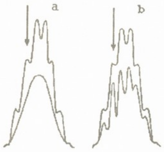

Fig.3. Schematical plot of pulse saturation of the EPR line of U^{3+} in CaF_2 at the orientation of the magnetic field H||[100]: a) pulse acting on the top fluorine HFS; b) pulse acting on the bottom.

When the pulse acted on the peak the whole structure disappeared and when it was between the peaks the whole structure was still better resolved (Fig.3). It happened in spring 1964 just before the conference on magnetic resonance in Krasnoyarsk. I hesitated before my report whether to tell or not about this effect and then decided to continue the investigation and first make it clear for myself. The fact that the effect observed is connected with hyperfine interaction was clear to me before.

In autumn of the same year I had distinct patterns of the burnt out spectrum of uranium in fluorite in main orientations of the magnetic field. At rotation of the magnetic field a strong angular dependence of the holes spectrum was observed. In the EPR line the structure disappeared but the pulse resolved it again. Numerous holes in EPR lines were also observed in other samples as well as in nonsinglecrystalline samples, for instance in the lines of a free radicals in the polyethylene irradiated in the nuclear reactor (Fig.4).

Going ahead, I should say that the observed hole spectrum or discrete saturation spectrum, as we call it, is a kind of a spin packet in the inhomogeneously broadened EPR lines. In this case it is impossible to burn out a single hole in the line. The entire hole spectrum appears together with the central hole and increases intensively with the pulse power or duration (Fig.5). The discrete saturation effect made everybody surprised but still remained an enigma - something strange.

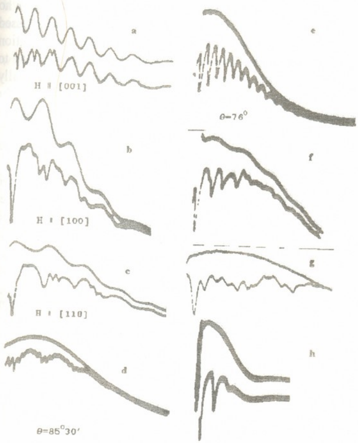

Fig.4. Oscillograms of DS spectra and the shape of the same portions of the line without saturating for tetragonal centers: a) - e) U^{3+} in CaF_2; f) Nd^{3+} in CaF_2; g) Yb^{3+} in CaF_2 H||[001]; h) free radicals in the polyethylene irradiated in the nuclear reactor.

Fig.5. Unperturbed shape of the EPR line (above) and DS spectra at two different levels of energy of the saturation pulse.

In 1966 in Tbilisi at the Institute of Physics a big conference on Plasma Physics was held. There I met Academician E.K.Zavoiski. He asked me to take him to my laboratory. "EPR is of more interest for me than plasma", - he said. As it is known, he had to stop his work on magnetic resonance. When he saw oscillograms of the spectra of the burnt-out holes he was very much surprised and insisted on publishing my work immediately even without explaining the cause of the effect.

I myself did not think of publishing the work in this way. The article on discrete saturation appeared in early 1967 [3]. In this article I intuitively guessed the main cause of DS - the change of the direction of the effective magnetic field acting as a nucleus surrounded by the magnetic center at an electron transition (Fig.6). In this case electron transitions with or without nuclear spin reorientation become of the same order of magnitude. There was only one step, perhaps half a step, to understand the mechanism of burning out of a hole spectrum, but it took me two years more. It is so clear and evident today, but that time nobody could take this step.

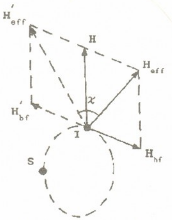

Fig.6. Schematical plot of the change of the direc-tion of the magnetic field on the nucleus at the electron transition.

For these two years I discussed this problem with such prominent scientists as Nicholas Bloembergen, who visited Tbilisi in 1967, and for a few days we discussed the causes of DS formation. In the same year a delegation of French physicists headed by Anatoli Abragam came to Tbilisi, they also discussed DS spectrum at a specially organized seminar. Discrete saturation was reported at the international symposium on free radical in Novosibirsk (1967) with an appeal to theorists to make an attempt to explain this phenomenon as well as in Grenoble in 1968 at the colloquies AMPERE.

During two years after the work had been published many scientists tried to explain this effect. And only after Grenoble I was suddenly enlightened and saw what should have been seen from the very beginning [4]. It is the most surprising in the history of DS technique.

Even today I cannot explain it. Why could nobody see what was so evident? Look how simple it is!

Let us consider the level scheme for eight equivalent

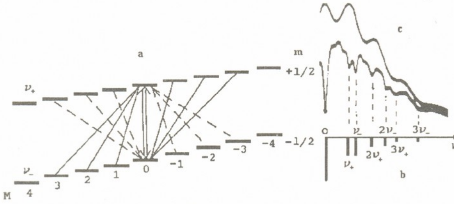

Fig.7. a) level scheme for the eight equivalent nuclei with I=1/2 surrounding the magnetic center with S=1/2; b) expected DS spectrum; c) experimentally observed spectrum for U^{3+} in CaF_2 H||[100].

nuclei with the spin I=1/2 surrounded by the magnetic center with effective spin S=1/2 (cubic centers in the fluorite type crystals at H||[100]* . In this scheme all kinds of electron transitions are allowed and that is important (Fig.7). Pulse saturation of the transition, indicated by an arrow, corresponds to the central hole. In this case the transitions indicated by solid lines correspond to the right hand side holes which at the same time are found in the state of partial saturation. Dashed lines indicate the transitions or holes left to the central hole.

And now look, how elegantly DS spectroscopy solves our initial task on explanation of the fluorine hyperfine structure of uranium in fluorite [5]. A pulse stroke

* For the magnetic centers of tetragonal symmetry in main orientations of the magnetic fields eight nuclei of the nearest neighbors are divided into two groups of four nuclei, but for the case under consideration - U^{3+} in CaF_2 approximation of the eight equivalent nuclei is quite acceptable.

like a stroke of a magic stick gives full information required for spectrum construction. It remains only to data enter into a computer. For this reason it is necessary to use the simple expression

$$2^{-9}[p(x^{s}+x^{-s}) + q(x^{r}+x^{-r})]^{8}(x^{\delta/2}+x^{-\delta/2}),$$

where

$$p = (4\nu_+ \nu_-)^{-1}[4\nu_F^2 - (\nu_+-\nu_-)^2];\ q = (4\nu_+ \nu_-)^{-1}[(\nu_+ + \nu_-)^2 - 4\nu_F^2];$$
$$s = 1/2(\nu_- - \nu_+);\ r = 1/2(\nu_- + \nu_+);$$

ν_+ and ν_- are splitting of upper and lower levels correspondently, δ -is splitting of levels by the ninth fluorine nucleus (Fig.1). This expression can be rewritten as a symmetrical polynom. The position of the spectral components is defined by degree of polynom's member and the intensity by the numerical coefficient in front of it. Each component is described by Gaussian form of line shape and its width is determined by the width of burning holes. This line is unusual. Each of its components has its own selection rules in order to change the total spin of eight fluorine nuclei surrounding the magnetic center: $\Delta M = 0, \pm1, ..., \pm8$, which gives 17 lines. 20 HFS lines (orientation [001]), are obtained by shift of this spectrum into three intervals, which is caused by the presence of an extra fluorine ion nucleus compensating an impurity charge according to B.Bleaney model of tetragonal center of uranium in fluorite (Fig. 8).

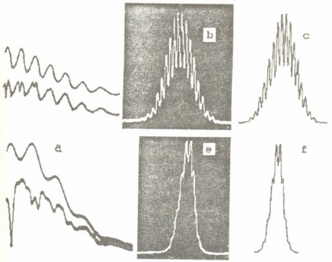

Fig.8. Tetragonal center U^{3+} in CaF_2: a) DS-spectrum; b) oscillogram of the fluorine HFS in orientation H||[001]; c) calculated spectrum (computer graphic); e)-g) the same for orientation H||[100].

Let us give another example of DS spectroscopy application. Recently, in studies of the EPR hole center or V_k - center in fluorite the DS method allowed to reveal a sodium ion near the V_k - center, whose presence had not been expected in the initial model [6]. DS spectroscopy allows to observe the sodium ion surrounded by the magnetic center directly on the oscillograph screen (Fig.9). Sample rotation allows to determine easily the direction and distance from it. Direction is determined by means of vanishing of DS holes and the distance by means of splitting in the holes spectrum. It looks like echoscopy, but it is a tiny magnetic nucleus near the magnetic center and not a calculus in the kidney.

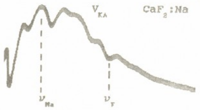

Fig.9. Oscillogram of DS-spectrum of V_k-centers in CaF_2:Na. Holes spectra from Na ions and F ions can be seen near the corresponding Larmor frequencies.

Radiofrequency discrete saturation (RFDS), as we call it, or DS-ENDOR represents the natural development of DS spectroscopy. It was suggested by Zevin and Brik in 1971 [7] and experimentally was discovered in 1972 by us [8,9].

When RF pulse follows the microwave pulse and radio frequency corresponds to splitting of electron levels by nuclei neighboring the magnetic center the resonance perturbation of DS spectrum was observed. As a result some DS holes must weaken and some new holes may appear. The application of RF field with the frequency ν_+ evokes redistribution of populations in the upper group of levels (Fig.10). Therefore it must cause the weakening of the system

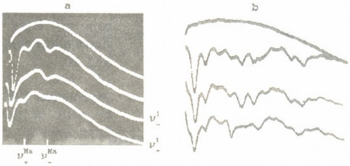

Fig.10. The signal of REDS: a) V_{KA}-center in the CaF_2:Na. The top oscillograme corresponds to the unperturbed shape of the EPR line. Lower - DS spectrum on the same line and action on resonance frequencies of the i-th nuclear surrounded by the magnetic center with are determind by weakening of the intensity of the first or second holes of Na ions; b) the same for more complicated DS-spectrum of Yb^{3+} in CaF_2.

of DS holes with the parameter ν_- and vice versa. So the DS-ENDOR method differs from the other ones with advantage of direct determination electron states. Saying figuratively in our case we get colour resonance frequencies whereas in the other ENDOR methods the frequencies are undistinguishable. Moreover, when different colour lines overlap because of separate receiving the frequencies they do not merge all together. Therefore resolution increases.

Let us consider application of DS-ENDOR method for V_k-center in CaF_2:Na. In Fig. 11 a is given angular dependence of sodium ion DS-ENDOR spectrum in the plane normal to the V_k-center axis. The agreement of the theory allowing for the low-symmetry quadrupole interaction and HFI with the experiment is illustrated by the plot.

In the second Fig.11 b the angular dependence for eight neighboring F nuclei divided into two nonequivalent groups surrounding Ca and Na ions is given. Another eight F nuclei are divided into three subgroups of nonequivalent nuclei (Fig.11c). The full information on HFI for all neighboring F ions is obtained, i.e. the HFI tensor components A are determined.

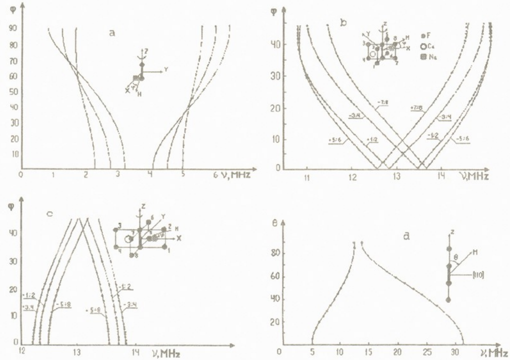

Fig.11. Angular dependence of RFDS spectra for V_k-center in CaF_2:Na: a) for Na^+ ion surrounded by V_k-center; b) for eight nearest ions of F^-, located by the ion of Ca^{2+} and ion of Na^+; c) for the next nearest eight ions of F^-; d) for the two F^--ions on the axis of molecule F_2^-.

In conclusion it should be noted that DS and DS-ENDOR pulse methods were suggested by us exactly 20 years ago at the 18-th Congress AMPERE in Nottingham, where Prof. Givi Khutsishvili was invited and made a review lecture [11]. These works formed the basis for the development of a new direction of EPR spectroscopy which now finds ever-widening application in the study of complex HFI in solids. Due to the development of this spectroscopy by Prof. A.Schweiger it becomes time domain and offers considerable promise for physics, chemistry and other fields.

I.Javakhishvili Tbilisi State University

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HISTORY OF SCIENCE

R.Shengelia

Treatment of Wounds in Georgian Traditional Medicine and New Aspects of Research in History of Medicine

Presented by Corr. Member of the Academy V.Mosidze, October 10, 1996

ABSTRACT. **Treatment of wounds in Georgian traditional medicine are discussed in the work. The material of about 500 manuscripts and more than 40 expeditions was analysed.**

The material is processed by computer. A special computer program was created and criteria and coefficient of reliability were determined. The latter is the basis for utilization/prohibition of traditional remedies. Strategy and policy of utilization of traditional remedies have been determined.

When analyzing all stages of the developmenl of medico-biological thought, when establishing every single stage, its regularity, cultural, political, economic phenomena, accompanying its development we come to the ideas changeable within time on correctness and universality of the way chosen by mankind. As a rule, revision of historical processes following a definite period of time is of speculative character, though it has a great importance when prognosing and planning future. While considering ancient medical manuscripts (anatomical and physiological treaties, medical books "karabadinis") and medical folklore (samples of folk medicine) we can distinguish two essential sides: 1) national, common to all mankind values - religion, philosophy, mythology, language, artistic thinking, which revealed in written monuments through mentality and views and in medical folklore they become evident in various rituals, legends, etc.; 2) the majority of remedies are produced from natural resources and thus their value and usage are ecologically justified. Their revival into everyday practice is inevitable which implies: assessment of efficiency, safety, quality, adaptation and application.

We have mainly investigated the second part of this two-sided problem, particularly, in nosology (a wound). The study of wound seems to be conditioned by the problems of secondary priority but in fact it is caused by recognition of problematic stagnation - a very characteristic trend of modern medicine. The wound treatment has not been improved during the past centuries and neither it has nowadays: a) the period of primary healing has not been reduced even in the world first class hospitals; b) the treatment of complicated wounds remains to be urgent. Any novelty in this direction is of great interest for surgeons all over the world.

When working on individual cases a number of general problems appeared calling for a response. Primarily we came across the fact that medical written monuments were not studied deeply. L.Kotetishvili, I.Abuladze, M.Saakashvili, M.Shengelia, A.Gelashvili, B.Rachvelishvili were pioneers and contributed a lot to the history of science. And still they could have some mistakes. Our work in this direction is to be regarded as appreciation and recognition of their contribution. This work includes: a) correct textual analyses of medical written monuments (historical, linguistic etc.); b) medical, philosophical analysis which are free of vulgar materialism; c) specificity of studies when investigating the epoch in relationship to foreign traditional medicine of the same period. Having compared historical facts we have established a fact of a considerable importance: the erroneous date of "Karabadini" by Kananeli (Book of Medicine) was shifted from XI century to the end of X century. And yet, our research focused on the question of specific practical importance of history of medicine. We believe this approach of science will take it's place granted to it since its origin. Due to some reasons it was lost in the course of time. The methods of Georgian traditional medicines, efficiency of remedies, their safety and quality that represent the coefficient of reliability - all this is a prerogative of a historian of medicine. A physician equipped with general historical methods, the one who has an experience and skills must do the following:

a) work on ancient medical written monuments;
b) identify natural resources and remedies;
c) establish the region of their spread;
d) find out parallels with foreign traditional medicine;
e) compare the results with the findings of modern medicine.

Then he has the right to give a primary evaluation of the methods of traditional medicine and remedies. Taking into account a number of parameters he can give recommendations for: a) practical use after minor tests; b) laboratory experimental studies at various levels; c) stating theoretical values; d) banning.

We posess WHO guidelines received in 1995 when the main conceptual system has been already formulated. We were pleased to find out a complete agreement with our positions though it's worth mentioning their free approach of WHO "Program of Traditional Medicine" on the method of traditional medicine and remedies. The authors of assessment of herbs give us a model of "Pharmokopea" where coefficient of reliability is limited by reference to ancient medical manuscripts. The confidence of these manuscripts is too high. We should notice that they contain all kinds of information: an exact identification of a plant, its spread, organoleptics, biochemical composition, physical and chemical tests. Special document is drawn on utilization of remedies, which cover the experience of developed and developing countries on production of remedies and on legislation and economic basis of their utilization.

The present work deals with wound treatment in Georgian traditional medicine and with *a priori* assessment of corresponding remedies, criteria of assessment and its system. First of all we should sum up the results of studies according to their quantitative index and determine the purposefulness and trends of tactics using remedies for wound treatment in ancient Georgian written monuments and folk medicine.

The propeties of the remedies were divided into two groups - basic and additional, which helped in case of radical treatment.The first group covers blood coagulation, antibacterial, antiinflamatory, wound closing, regerenation stimulating and analgetic remedies. The second group covers: biostimulators, tranquilizers-fever, sudorific, diuretics and swelling resolve remedies. This approach was used both towards ancient classical and folk medicine and it has been improved by us when assessing remedies of respiratory system.

On the basis of information analyses being at our disposal, every remedy is indexed by its characteristic in literature: basic or additional. Proceeding from the above we can judge about wound treatment, tactics of treatment, about the significance of the source in general and the epoch it belonged to.

We have studied 4 written manuscripts (in total they exceed 500). Kananeli's "Ustsoro Karabadini", "Tsigni Saakimo" (Medical Book) by Khojasvili - XII c; Zaza Panaskerteli-Tsitsishvili's "Samkurnalo Tsigni Karabadini" XV c. and David Batonishvili's "Iadigar Daudi" XVI c. They all are original high-level scientific works for their period.

Wound treatment occupies a significant place in Georgian traditional medicine but proceeding from the extremely high-level surgery in Georgian written medical monuments we believe in a possible existence of a special surgical monograph which has not come to our days.

Ancient Georgian surgeons distinguished wounds by their shapes (incised, stab, arrow induced, etc.). intensity of infections and bleeding, lingering, topographic location threat for life. They described a clinical picture, a patient's general state.

The treatment which was based on humoral pathological theory in addition to physicomechanical one (suture, draining tamponi, removal of foreign bodies by special instrument, excision of necrotic tissues) was strictly purposeful. The main purpose of treatment was to achieve blood coagulation, wound closure, regeneration and was also directed to promotion of antibacterial, antiinflamatory and analgetic effects.

Georgian surgeons used symptomatic-biostimulating, tranquilizing, antifever, sudorific, diuretic and swelling resolve means. These means are mainly combined remedies.

In the Georgian written monuments of medieval age the issues of diet which was also based on humoral theory, played a significant role in wound treatment and was directed to the health improvement.

The Georgian written monuments touch upon issues of ethic, based on a surgeon's high professionalism and honesty.

Georgian folk medicine is closely linked with ancient Georgian classical medicine, revealed in the mutual enrichment within centuries. Georgian folk medicine makes use of 138 herbs, among them 122 are used in modern medicine. We failed to identify 11 herbs owing to inadequate entries made during expeditions. Georgian folk medicine also uses 6O remedies of animal and 33 remedies of mineral origin.The natural medicines include combined and monocomponent remedies.

Contrary to classical medicine, Georgian folk medicine makes use of herbs characterized by a number of properties. The question is how strong these properties are, how strong are active ingredients accumulated in the plant. We can *a priori* consider these facts proceeding from the intensity of specific use of the plant, their study and literature. Hence, two tendencies became exident: vertical and horizontal ones. The first was called extensive characteristic of properties while the second one - intensive characteristic of properties. The rating of remedies is the sum of the cardinal parameters computer-aided storing, systematization.

Advanced computer aided method of information storing, systematisation and evolution was used to achieve this goal.

A special original program with mathematical insurance was created. A great amount of information was put in computer: 74 reference books, atlas, vocabulary and the variation line of " coefficient of reliability " with functional limits was formulated. Algorithm was determined by 12 data system. Among them two are informational and ten are parametric.

Informational data are names of remedies in Georgian, Latin and in IO foreign languages. Parametric data: spread, utilization, form of remedies, perfectness of prescription, level of study, known pharmacological effects, their use in modern medicine.

The mentioned informational and parameter system gave us a strictly determined individual pattern of every single remedy rating of which was numerically expressed in "coefficient of reliability". The upper level of coefficient is theoretically infinite.

In fact, in individual cases of wound treatment the optimal limit of coefficient of reliability can be determined from 3 500 to 4 700. This range enables us not only permit producton of herbal medicine but also make recommendations on their expected high efficiency.

Remedies which are in the area from 2 200 to 3 300 can be also used but the producer himself bears a moral, professional and juridical responsibility for their efficiency.

To our mind if the coefficient of reliability is under 2 000 a laboratory and experimental tests and conform of their efficiency and safety we can proceed the usage of remedies. In case of combined remedies it will come to arithmetic summing up.

Some well known remedies, for example Valerian, have unusually a low coefficient. It should be taken into consideration that the mentioned remedies are assessed for the purpose of a specific wound treatment and the mechanism of the fact mentioned becomes evident.

Hence it becomes necessary to create two types of computer models: particular and general. The latter is to be a universal index of natural materials. For time being we have avoided the assessment of animal and mineral natural material owing to the fact that with several exceptions the material has not been studied. Though we have created a computer card. We have also created a scheme of use of traditional medicine, where we tried to establish the principal sources, stages and structure which provide an efficient application of tradidonal medicine.

We are far from claiming that computerized study of natural material gives us an exact picture. Just the opposite, the method must enhance an individual approach for each type of material.

Nothing can substitute an investigator's intuition based on his knowledge. In addition to planned methodical study of traditional medicine and its application, this field is fraught with unexpected discoveries. .

Tbilisi State Medical University.

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INDEX 76181