ISSN—0132—1447

BULLETIN
OF THE GEORGIAN ACADEMY OF SCIENCES

საქართველოს მეცნიერებათა აკადემიის
მოამბე

154

№ 3

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

BULLETIN
OF THE GEORGIAN ACADEMY OF SCIENCES

საქართველოს მეცნიერებათა აკადემიის
მოამბე

Volume 154

№ 3

1996

თბილისი * TBILISI

The Journal is founded in 1940

BULLETIN

OF THE GEORGIAN ACADEMY OF SCIENCES

is a scientific journal, issued bimonthly in
Georgian and English languages

Editorial Office:
Georgian Academy of Sciences
52, Rustaveli Avenue,
Tbilisi, 380008,
Republic of Georgia

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

Dear Colleagues,

The "Bulletin of the Georgian Academy of Sciences" is the leading scientific journal in Georgia. Since 1996 it's going to be published in Georgian and English.

The Journal publishes mainly the articles featuring the results of research carried out in scientific institutions of Georgia.

Publication of the Journal in English will make it possible to a foreign scientific community to get to know the up-to-date achievements in Georgian Science and at the same time, it'll give the chance to draw the scientists from the other countries to cooperate with our Journal.

We hope that you, having subscribed to our Journal, will assist in overcoming the difficulties, connected with its publication. We also hope that you'll use your authority and influence for popularization of the "Bulletin".

A. Tavkhe

Academician Albert N. Tavkhelidze

Editor in Chief
The President of the Georgian
Academy of Sciences

CONTENTS

MATHEMATICS

CYBERNETICS

PHYSICS

GEOPHYSICS

ORGANIC CHEMISTRY

HUMAN AND ANIMAL PHYSIOLOGY

BIOPHYSICS

BIOCHEMISTRY

CYTOLOGY

PALEOBIOLOGY

LINGUISTICS

HISTORY

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

Corr. Member of the Academy N.Berikashvili, D.Makalatia

The Multiplicative Version of Twisted Tensor Product Theorem

Presented December 11, 1995

ABSTRACT. **A version of Brown's twisted tensor product theorem is given where the model carries the multiplicative structure.**

1. **Introduction**. Brown's twisted tensor product theorem [1] alters in tensor product of chain complexes of base and fiber of fibration the differential in such a way, that the resulting chain complex is chain equivalent with the chain complex of fibration. But in this construction the multiplicative structure of cohomology ring is neglected. The first author in case of fiber Eilenberg-MacLane space removed this non convenience [2,3]. In this note we assert that the result is valid for arbitrary fibers. The method of approach is that of [2,3]. We confine ourselves with the case of fiber bundles.

2. **Case of homology**. Let $F \to E \to B$ be a fiber bundle with the structural group G and let $G \to \overline{E} \to B$ be the associated principal fibration. Let $S(X)$ and $Q(X)$ denote singular and cubical singular complexes of space X respectively, $C_*(X)$ and $C_*^{cu}(X)$ denote the singular chain complex and the cubical chain complex of space X respectively. Recall that $C_*^{cu}(X) = \widetilde{C}_*^{cu}(X)/\overline{C}_*^{cu}(X)$ is the chain complex of all cubical chains factored by the subcomplex generated by degenerated cubes. Then $C_*^{cu}(G)$ is a graded differential algebra with the Pontrjagin product coming from group product

$G \times G \to G$, the group $C_*^{cu}(F)$ is the graded module over the graded differential algebra $C_*^{cu}(G)$ coming from action of group G on the space F, $G \times F \to F$. Consider the bigraded differential module $A(B, G) = C^*(B, C_*^{cu}(G))$ singular cochains of B with the coefficients of cubical chains of G. It is a bigraded differential algebra with $\cup$-product of singular cochains of B and Pontrjagin product of coefficients. Let $h^I \in C^*(B, C_*^{cu}(G))$, $h^I = h^{2,-1} + h^{3,-2} + h^{4,-3} + \ldots$, be twisting cochain, i.e. such that $d_A h^I = h^I \cup h^I$. We have the equality

$$C_*(B) \otimes C_*^{cu}(F) = C_*(B, C_*^{cu}(F))$$

Denote it by $M(B, F)$. It is a bigraded differential module over the bigraded differential algebra $A(B, G)$, the structure defined by $\cap$ -product in B and module structure of $C_*^{cu}(F)$. Then the element h^I defines the new differential in this bigraded differential module, the twisted differential, by the formula

$$d_{h^I} x = dx + h^I \cap x.$$

The differential module M with this differential is denoted by $C_*(B) \otimes_{h^I} C_*^{cu}(F)$.

Theorem 1. *There is a rule which assigns to every principal G-fibration* $\overline{E}$, $G \to \overline{E} \to B$, *a twisting cochain* $h^{1}_{\overline{E}} \in C^{*}(B, C^{cu}_{*}(G))$ *such that:* (*i*) *it is functorial, that means for* $f\colon B' \to B$ *and* $G \to \overline{E}' \to B'$ *the induced fibration one has* $h^{1}_{\overline{E}'} = f^{*}(h^{1}_{\overline{E}})$ (*ii*) *if E is an associated bundle with the fiber F then there is a chain map* $t\colon C_{*}(B) \otimes_{h^{1}_{\overline{E}}} C^{cu}_{*}(F) \to C^{cu}_{*}(E)$ *inducing isomorphism of homologies.*

This theorem is essentially Brown's theorem modified for cubical chain complexes.

Sketch of proof. The proof is based on the notion of quasi-cubical set which carries degeneracy operators only partially: (*i*) the n-cubes are indexed by bidegree, $\sigma^{p,q}$, $p+q = n$; (*ii*) there are face operators d^{ε}_{j}, $j = 1, 2, \dots, (p+q)$, $\varepsilon = 0, 1$, and degeneracy operators s_{p+j}, $j = 1, 2, \dots, q$, subject to the ordinary equalities. The cube $d^{\varepsilon}_{j}(\sigma^{p,q})$ has bidegree $(p-1, q)$ if $j \le p$ and $(p, q-1)$ if $p < j \le (p+q)$. The cube $s_{p+j}(\sigma^{p,q})$ has bidegree $(p, q+1)$. (*iii*) for fixed p it is an ordinary cubical set with the dimension of cube $\sigma^{p,q}$ to be q, the face operators for $j \le q$ to be d^{ε}_{p+j} and degeneracy operators s_j, $j = 1, 2, \dots, q$, to be s_{p+j}. The proof proceeds in three steps:

1) Constructing a map $\kappa\colon S_{*}(B) \to Q_{*}(B)$ subject to properties:

(1) $\dim \kappa(\sigma^{m+1}) = m$;

(2) $\kappa(\sigma^{1}) = e$ unity of group G;

(3) $d^{1}_{i}(\kappa(\sigma^{m+1})) = \kappa(\sigma^{m+1}_{i})$, $1 \le i \le m$;

$d^{0}_{i}(\kappa(\sigma^{m+1})) = \kappa(\sigma^{i}_{1})\kappa(\sigma^{m+1-i}_{2})$, $\sigma^{i}_{1} \cup \sigma^{m+1-i}_{2} = \sigma^{m+1}$, $1 \le i \le m$.

2) Converting the cartesian product $S(B) \times Q(F)$ in quasi-cubical complex by

$d^{\varepsilon}_{p+i}(\sigma^{p} \otimes \tau^{q}) = \sigma^{p} \otimes d^{\varepsilon}_{i}\tau^{q}$, $1 \le i \le q$, $\varepsilon = 0, 1$;

$d^{1}_{j}(\sigma^{p} \otimes \tau^{q}) = \sigma^{p}_{j-1} \otimes \tau^{q}$ $1 \le j \le p$;

$d^{0}_{j}(\sigma^{p} \otimes \tau^{q}) = \sigma^{j-1}_{1} \otimes \kappa(\sigma^{p-j+1}_{2})\tau^{q}$, $1 \le j \le p$, $\sigma^{j-1}_{1} \cup \sigma^{p-j+1}_{2} = \sigma^{p}$;

$s_{p+i}(\sigma^{p} \otimes \tau^{q}) = \sigma^{p} \otimes s_{i}\tau^{q}$, $1 \le i \le q$.

Resulting quasi-cubical complex, the twisted product, denote $S(B) \times_{\kappa} Q(F)$.

3) Constructing a map

$$S(B) \times_{\kappa} Q(F) \to Q(E)$$

preserving the face and degeneracy operators and inducing isomorphism of homology.

4) Observing that the reduced chain complex of quasi-cubical complex $S(B) \times_{\kappa} Q(F)$ is exactly $C_{*}(B) \otimes_{h^{1}} C^{cu}_{*}(F)$ with h defined by $h(\sigma^{m+1}) = \kappa(\sigma^{m+1})$.

3. **Case of cohomology**. Recall that for the topological space X the cubical cochain complex $C^{*}(X, \Lambda)$ with the coefficients in the module Λ are such cochains which have value zero on each degenerate cube of X. If A is ring then Serre $\cup$-product is defined as follows [4]. For the subset $K = (i_1 < i_2 < i_3 \dots < i_k) \subset (1, 2, 3, \dots, n)$ and σ^{n} the n-cube of X we introduce the notation

$$d^{\varepsilon}_{K}\sigma^{n} = d^{\varepsilon}_{i_1} d^{\varepsilon}_{i_2} d^{\varepsilon}_{i_3} \dots d^{\varepsilon}_{i_k}(\sigma^{n}),\ \varepsilon = 0, 1.$$

Let $c^p \in C^p(X, \Lambda)$, $c^q \in C^q(X, \Lambda)$ be the normalized p and q-cochains of $Q(X)$ with coefficients in a commutative ring Λ. Define the product $c^{p+q} = c^p \cup c^q$ by

$$c^{p+q}(\tau^{p+q}) = \sum_{(H,K)} (-1)^{\alpha(H,K)} c^p(d_K^0(\tau^{p+q}))c^q(d_H^1(\tau^{p+q}))$$

where (H, K) is the decomposition of $\{1, 2, 3, \ldots, (p+q)\}$ into two disjoint subsets. This formula defines product in arbitrary cubical complex, in particular, in complex $S(B) \times_\kappa Q(F)$. Hence the cochain complex

$$C^*(S(B) \times_\kappa Q(F), \Lambda) = hom\,(C_*(B) \otimes_{h_E} C_*^{cu}(F), \Lambda)$$

carries the multiplicative structure. Below we assume that

$$C^*(S(B) \times_\kappa Q(F), \Lambda) = hom\,(C_*(B) \otimes_{h_E} C_*^{cu}(F), \Lambda)$$

is considered with this multiplicative structure. Hence we have defined an algebraic model of bundle which carries a multiplicative structure as well. In sketch of proof of theorem 1 we noted that t: $S(B) \times_\kappa Q(G) \to Q(E)$ preserves face and degeneracy operators, hence we have

Theorem 2. *Induced by the chain map t of theorem 1 the map of cochain complexes* $t^*: C_{cu}^*(E, \Lambda) \to hom\,(C_*(B) \otimes_h C_*^{cu}(F), \Lambda)$ *preserves the multiplicative structure.*

The research described in this publication was made possible in part by Grant NoRVCOOO from the International Science Foundation.

A.Razmadze Mathematical Institute
Georgian Academy of Sciences

REFERENCES

1. *E.H.Brown*. Ann. of Math. **69**, 223-246, 1959.
2. *N.Berikashvili*. Bull. Georg. Acad.Sci., **152**, 3, 1995.
3. *N.Berikashvili*. Georgian Mathematical J., **3**, 1, 27-48, 1996.
4. *J.P.Serre*. Ann. of Math., **54**, 429-505, 1954.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

Z.Samsonia

On Quasiconformal Mapping of Close Domains

Presented by Academician I.Kiguradze, September 6, 1996

ABSTRACT. **The problem of quasiconformal mapping of the following Beltrami equation is set and studied for close domains in a definite sense.**

Let the coefficient $q(z)$ of Beltrami equation

$$W_{\bar{z}} = q(z) W_z \tag{1}$$

be given a finite simply connected domain G_0 and belong to the class of functions $C_\gamma(\overline{G_0})$ $(0 < \gamma \leq 1)$; besides $z = 0 \in G_0$. We denote by $\widetilde{W}(z)$ ($\widetilde{W}(\infty) = \infty$, $z^{-1}\widetilde{W}(z) \to 1$, as $|z| \to \infty$) the so-called principal homeomorphism of the whole complex plane E (of the equation (1)) constructed according to I.Vekua's scheme [1] under the condition that $q(z) = 0$ outside $\overline{G_0}$.

Let's consider on the complex plane simply connected domain G $(G \subseteq G_0$, $0 \in G)$,bounded by Liapunov's contour $\Gamma = \partial G$ $(\Gamma \in C^1_\alpha$, $0 < \alpha \leq 1)$, whose parametric equation is

$$g = g(\tau),\ (0 \leq \tau \leq 2\pi).$$

It is shown [2] that a function

$$W(z) = [\,\widetilde{W}(z) - \widetilde{W}(0)]\exp\left(\frac{1}{\pi i}\int_\Gamma \frac{\nu(t)\,d\widetilde{W}}{\widetilde{W}(t) - \widetilde{W}(z)} + iC(\nu)\right), \tag{2}$$

where $d\widetilde{W} = \widetilde{W}_t\,dt + \widetilde{W}_{\bar{t}}\,d\bar{t}$ satisfies the equation (1) and univalently maps G to unit disk so, that $W(0) = 0$, $W(z_1) = 1$, $z_1 \in \Gamma$ $(z_1 > 0)$, if $\nu(t)$ is a solution of Fredholm integral equation

$$\nu(t) + \operatorname{Re}\frac{1}{\pi i}\int_\Gamma \nu(\tau)\frac{d\widetilde{W}}{\widetilde{W}(\tau) - \widetilde{W}(t)} = -\ln\left|\widetilde{W}(t) - \widetilde{W}(0)\right| \tag{3}$$

and

$$C = -\arg[\,\widetilde{W}(z_1) - \widetilde{W}(0)] - \frac{1}{\pi}\int_\Gamma \nu(t)\,d\ln\left|\widetilde{W}(\tau) - \widetilde{W}(z_1)\right| \tag{4}$$

The equation (3) has a unique solution which belongs to Hölder class.

According to [3] we can say that $\widetilde{W}_z, \widetilde{W}_{\bar{z}} \in C_{\gamma_0}(\overline{G})$, where $0 < \gamma_0 \leq \min\{\alpha, \gamma\}$.

Now let us consider a domain $\widetilde{G}$ $(\widetilde{G} \subset G_0)$ of the type G, let the parametric equation of its boundary $\partial\widetilde{G} = \widetilde{\Gamma} \in C^1_\alpha$ be

$$g = \widetilde{g}(\tau),\ (0 \leq \tau \leq 2\pi).$$

We can use the scheme (2)-(4) to construct the function $\widetilde{W}_{\widetilde{\nu}}(z)$ ($\widetilde{W}_{\widetilde{\nu}}(0) = 0$, $\widetilde{W}_{\widetilde{\nu}}(\widetilde{z}_1) = 1$, $\widetilde{z}_1 \in \widetilde{\Gamma}$, $\widetilde{z}_1 > 0$) which maps are quasiconformally to the unit disk $\widetilde{G}$.

Definition. We call the domain $\widetilde{G}$ ε-close ($\varepsilon > 0$) to G, of the conditions are fulfilled

$$\max_{\tau} |g(\tau) - \widetilde{g}(\tau)| \leq \varepsilon, \quad \|g'(\tau) - \widetilde{g}'(\tau)\|_{C_\alpha} \leq \varepsilon \tag{5}$$

For each $\varepsilon > 0$ the infinite set of domains ε-close to G is defined which we denote by G_ε.

By passing to the plane of homeomorphism $\widetilde{W}(z)$ the integral equation (2) and the integral equation constructed for the domain $\widetilde{G}$ close to G become:

$$\mu(\xi_0) + \operatorname{Re} \frac{1}{\pi i} \int_{\widetilde{W}(\Gamma)} \frac{\mu(\xi) d\xi}{\xi - \xi_0} = -\ln|\xi_0 - W_0|, \ \xi_0 \in \widetilde{W}(\Gamma), \tag{2'}$$

$$\widetilde{\mu}(\xi_0) + \operatorname{Re} \frac{1}{\pi i} \int_{\widetilde{W}(\widetilde{\Gamma})} \frac{\widetilde{\mu}(\xi) d\xi}{\xi - \xi_0} = -\ln|\xi_0 - W_0|, \ \xi_0 \in \widetilde{W}(\widetilde{\Gamma}), \tag{$\widetilde{2}'$}$$

where

$$\mu(\xi_0) = \nu\left[\widetilde{W}(g(\tau_0))\right], \ \widetilde{\mu}(\xi_0) = \widetilde{\nu}\left[\widetilde{W}(\widetilde{g}(\tau_0))\right], \ \tau_0 \in [0, 2\pi],$$

$$W_0 = \widetilde{W}(0), \ \xi = \widetilde{W}[g(\tau)] \text{ in (2')}, \ \xi = \widetilde{W}[\widetilde{g}(\tau)] \text{ in } (\widetilde{2}').$$

The following statement is true.

Theorem. *If the domains $G \subseteq G_0$ and $\widetilde{G} \in G_\varepsilon$ ($\widetilde{G} \subset G_0$) whose boundaries belong to the class C^1_α ($0 < \alpha \leq 1$) are ε-close to each other, then for each $\varepsilon \in [0, \varepsilon_1]$ the inequality*

$$\|\mu(\xi) - \widetilde{\mu}(\xi)\|_{C_{\gamma_0 - \eta - \beta}} \leq C_0^*(\widetilde{W}; G; \beta) \|\mu\|_{C_{\gamma_0 - \eta - \beta}} \varepsilon^{\eta} \tag{6}$$

holds, where $\mu(\xi) = \nu[\widetilde{W}^{-1}(g(\tau))]$, $\widetilde{\mu}(\xi) = \widetilde{\nu}\left[\widetilde{W}^{-1}(\widetilde{g}(\tau))\right]$ are the solutions of integral equations (2') *and* ($\widetilde{2}'$)*, $0 < \gamma_0 \leq \min\{\alpha, \gamma\}$, β is arbitrary positive number less than $\gamma_0 - \eta$, and $\eta \in (0; \gamma_0)$. The constant C_0^* and small number $\varepsilon_1 > 0$ may be entirely defined by giving the initial domain G and homeomorphism $\widetilde{W}(z)$.*

This theorem gives a possibility to define the approximation of the function $W = W_{\widetilde{\nu}}(z)$ which maps quasiconformally an arbitrary domain $\widetilde{G} \in G_\varepsilon$ ($0 < \varepsilon \leq \varepsilon_1$) to the disk $|W| < 1$ by formula

$$W_\nu(z) = [\widetilde{W}(z) - \widetilde{W}(0)] \exp\left(\frac{1}{\pi i} \int_{\Gamma} \frac{\nu(t) d\widetilde{W}}{\widetilde{W}(t) - \widetilde{W}(z)} + i\widetilde{C}(\nu)\right).$$

The estimation of the error $|W_{\tilde{\nu}}(z)-W_{\nu}(z)|$ in $\tilde{G}$ is possible using (6) and the method of proof of the Plemel-Privalov's theorem [4].The error has the same order $O(\varepsilon^{\eta})$.

N.Muskhelishvili Institute of Computational
Mathematics
Georgian Academy of Sciences

REFERENCES

1. *I.Vekua.* The Generalized Analytic Functions, M., 1953.
2. *D.Kveselava, Z.Samsonia.* In: "Metrical Questions of Theory of Functions. Kiev, 1980.
3. *G.Mandjavidze.* Papers of the Math. Inst. Tbilisi, XXXIII, 1967.
4. *N.Muskhelishvili.* Singular Integral Equations. M., 1962.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

M.Chkhartishvili

Notes on the Convergence of the Scheme of Increased Accuracy for one Mixed Boundary Value Problem

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

ABSTRACT. **The paper deals with the estimate of convergence of difference scheme for Poisson equation with mixed boundary value conditions. The estimate of convergence compatible with the smoothness of the exact solution was known for $2 < S \le 4.5$. In this paper the indicated estimate is obtained for $4.5 \le S \le 5$.**

This paper investigates the convergence of the difference scheme [1] approximating Poisson equation with mixed boundary value conditions:

$$\frac{\partial^2 u}{\partial x_1^2}+\frac{\partial^2 u}{\partial x_2^2}=-f(x), \quad x\in\Omega, \tag{1}$$

$$\frac{\partial u}{\partial x_1}=-g(x), \quad x\in\Gamma_{-1}; \quad u(x)=0, \quad x\in\Gamma\setminus\Gamma_{-1}. \tag{2}$$

Here $\Omega=\{(x_1,x_2)\colon 0<x_\alpha<l_\alpha,\ \alpha=1,2\}$ is a rectangle with the boundary Γ, Γ_{-1} is its left vertical side.

The estimate of convergence rate

$$\|y-u\|_{W_2^1(\omega)}\le M|h|^{S-1}\|u\|_{W_2^S(\Omega)}, \quad 2<S<4.5 \tag{3}$$

has been established in [1] for the corresponding difference scheme.

For more smooth solutions, i.e. when $4.5 \le S \le 5$, the loss of half a convergence rate takes place.

On the other hand, it is expected that the estimate (3) will be valid for a wider set of index S. Namely, the validity of inequality (3) for an arbitrary parameter S from the interval (2;5] is shown in this paper.

To get more precise a priori estimate (in the sense of norm compatibility $\|\cdot\|_{W_2^{S-1/2}(\Gamma)}\le M\|\cdot\|_{W_2^S(\Omega)}$) as it is in [2,3,4] the aforementioned gap can be removed.

Let's consider the difference scheme from [1].

In $\overline{\Omega}=\Omega\cup\Gamma$ we define the grid $\overline{\omega}=\overline{\omega}_1\times\overline{\omega}_2$, where

$$\overline{\omega}_\alpha=\{x_\alpha=i_\alpha h_\alpha\colon\ i_\alpha=0,1,\dots,N_\alpha;\ h_\alpha=l_\alpha/N_\alpha\}, \quad \alpha=1,2.$$

Let

$$\omega_\alpha^+=\overline{\omega}_\alpha\cap(0;l_\alpha], \quad {}^+\omega_\alpha=\overline{\omega}_\alpha\cap[0;l_\alpha), \quad \omega_\alpha=\overline{\omega}_\alpha\cap(0;l_\alpha), \quad \alpha=1,2; \quad \gamma=\overline{\omega}\setminus\omega,$$

$$\gamma_{-1}=\{x=(0;x_2)\colon x_2\in\omega_2\},\ |h|^2=h_1^2+h_2^2,\ \hbar_1=h_1/2,\ \ x\in\gamma_{-1},\ \hbar_1=h_1,\ x\notin\gamma_{-1}.$$

Let also

$$(y,v)=\sum_{{}^+\omega_1\times\omega_2}\hbar_1 h_2 yv, \quad \|y\|=(y,y)^{1/2},$$

$$\|\nabla y\|^2=\|y\|_{W_2^1(\omega)}=\sum_{\omega_1^+\times\omega_2} h_1h_2y_{\bar{x}_1}^2+\sum_{\omega_1\times\omega_2^+}\hbar_1h_2y_{\bar{x}_2}^2,$$

$$T_\alpha u(x)=\frac{1}{h_\alpha^2}\int_{x_\alpha-h_\alpha}^{x_\alpha+h_\alpha}\left(h_\alpha-\left|\xi_\alpha-x_\alpha\right|\right)u(\xi_1,\xi_2)d\xi_\alpha,\ \xi_{3-\alpha}=x_{3-\alpha},\ \alpha=1,2,$$

$$T_1^+u(x_1,x_2)=\frac{2}{h_1^2}\int_{x_1}^{x_1+h_1}(h_1+x_1-\xi_1)u(\xi_1,x_2)d\xi_1.$$

The boundary value problem (1), (2) is approximated by the difference scheme [1].

$$\Lambda y\equiv(\Lambda_1+\Lambda_2+\frac{h_1^2+h_2^2}{12}\Lambda_1\Lambda_2)y=-\varphi(x),\quad x\in\omega\cup\gamma_{-1},$$

$$y(x)=0,\quad x\in\gamma\backslash\gamma_{-1},\tag{4}$$

where

$$\Lambda_1y=\begin{cases}y_{\bar{x}_1x_1}, & x\in\omega,\\ \frac{2}{h_1}y_{x_1}, & x\in\gamma_{-1},\end{cases}\qquad \Lambda_2y=y_{\bar{x}_2x_2},$$

$$\varphi(x)=\begin{cases}T_1T_2f, & x\in\omega,\\ T_1^+T_2f+\frac{2}{h_1}T_2g-\frac{h_1}{6}g_{\bar{x}_2x_2}, & x\in\gamma_{-1}.\end{cases}$$

3. The error $z=y-u$ satisfies the problem

$$\Lambda z=\psi,\quad x\in\omega\cup\gamma_{-1},$$

$$z(x)=0,\quad x\in\gamma\backslash\gamma_{-1},\tag{5}$$

where

$$\psi=\begin{cases}\eta_{1\bar{x}_1x_1}+\eta_{2\bar{x}_2x_2}, & x\in\omega,\\ \frac{2}{h_1}(\eta_{1x_1}+\eta_{x_2}), & x\in\gamma_{-1},\end{cases}$$

$$\eta_{3-\alpha}(x)=T_\alpha u-u-\frac{h_\alpha^2}{12}T_\alpha\frac{\partial^2u}{\partial x_\alpha^2},\quad \alpha=1,2,$$

$$\eta=\frac{h_1}{2}\left(T_1^+u-u-\frac{h_1}{3}\frac{\partial u}{\partial x_1}-\frac{h_1^2}{12}T_1^+\frac{\partial^2u}{\partial x_1^2}\right)_{\bar{x}_2}.$$

From (5) it follows that

$$(-\Lambda z,z)=\sum_{\omega_1^+\times\omega_2}h_1h_2\eta_{1\bar{x}_1}z_{\bar{x}_1}+\sum_{\omega_1\times\omega_2^+}h_1h_2\eta_{2\bar{x}_2}z_{\bar{x}_2}-\sum_{\gamma_{-1}}h_2\eta_{x_2}z.\tag{6}$$

We apply the Cauchy-Buniakovski inequality to the first and second sums of the right side and for the estimate of the last sum we use the inequality ([4])

$$\left|\sum_{\gamma_{-1}} h_2\eta_{x_2}z\right| \le \|\nabla z\|\left(\left(\sum_{\omega_1\times\omega_2^+} h_1h_2\eta_{\bar{x}_1}^2\right)^{1/2}+\left(\sum_{{}^+\omega_1\times\omega_2} h_1h_2\eta_{x_2}^2\right)^{1/2}\right),$$

then by virtue of the inequality

$$(-\Lambda z, z) \ge \frac{2}{3}\|\nabla z\|^2,$$

we obtain:

$$\|\nabla z\| \le \frac{3}{2}\left(\|\eta_{1\bar{x}_1}\|_{\omega_1^+\times\omega_2}+\|\eta_{2\bar{x}_2}\|_{\omega_1\times\omega_2^+}+\|\eta_{\bar{x}_1}\|_{\omega_1\times\omega_2^+}+\|\eta_{x_2}\|_{{}^+\omega_1\times\omega_2}\right). \tag{7}$$

The linear functionals $\eta_{\alpha\bar{x}_\alpha}$ $(\alpha=1,2)$, $\eta_{\bar{x}_1}$, η_{x_2} vanish on polynomials of the forth power and are limited on functions from the class $W_2^S(\Omega)$, $S>2$.

Consequently, using the technique developed in [1], it can be showed, that their modules are bounded from above by the value $M|h|^{S-1}(h_1h_2)^{-1/2}|u|_{W_2^s(e)}$, $S>2$. e – is an elementary cell on grid $\bar{\omega}$. It leads to an estimate of the norms on the right handside of the inequality (7) by $M|h|^{S-1}|u|_{W_2^s(\Omega)}$, $2<S\le5$, and therefore, following statement on the accuracy of the difference scheme (4) is valid.

Theorem. *Let for the problem (1), (2) conditions, providing the belonging of the solution $u(x)$ to the class $W_2^S(\Omega)$, $2<S\le5$ are fulfilled. Then the rate of the convergence of the difference scheme (4) is characterised by estimate*

$$\|y-u\|_{W_2^1(\omega)} \le M|h|^{S-1}|u|_{W_2^s(\Omega)}, \quad S\in(2;5],$$

where the constant M does not depend on $|h|$ and $u(x)$.

Georgian Technical University

REFERENCES

1. *A.A.Samarski, R.D.Lazarov, V.L.Makarov.* Difference Schemes for Differential Equations with Generalized Solutions. Moscow, 1987 (Russian).
2. *G.K.Berikelashvili.* Bull. Acad. Sci. Georgian SSR, **118**, 2, 1985, 285-288.
3. *M.D.Chkhartishvili, G.K.Berikelashvili.* Bull. Acad. Sci. Georgia, **148**, 2, 1993, 180-184.
4. *G.K. Berikelashvili, M.D. Chkhartishvili.* Bull. Acad. Sci. Georgia, **153**, 2, 1996, 195-198.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

M.Usanetashvili

An Oblique Derivative Problem for a Second Order Elliptic Equation with a Parabolic Degeneration on a Part of Boundary

Presented by Academician I.Kiguradze, July 8, 1996

ABSTRACT. **The oblique derivative problem is considered for a second order elliptic differential equation which is degenerated on some part of a boundary. The uniqueness and existence of theorems are proved.**

Let us consider the equation

$$L(u) = AU_{xx} + 2BU_{xy} + CU_{yy} + aU_x + bU_y + cU = F, \tag{1}$$

in the bounded, simply connected domain D, where

$$A, B, C \in C^{2,\alpha}(\overline{D}),\ a,\ b \in C^{1,\alpha}(\overline{D}),\ c \in C^{0,\alpha}(\overline{D}),\ F \in C(D),\ 0 < \alpha < 1. \tag{2}$$

$$c \le c_0 = \text{const} < 0.$$

Let $\partial D = \Gamma_1 \cup \Gamma_2 \cup P_1 \cup P_2$, $\Gamma_1 \cap \Gamma_2 = \varnothing$, where $\overline{\Gamma}_1$ and $\overline{\Gamma}_2$ are open areas with the end points P_1 and P_2; $\overline{\Gamma}_1$ and $\overline{\Gamma}_2$ belong to $C^{3,\alpha}$ class.

We suppose that the equation (1) is elliptic in the domain $DU\Gamma_1$ and it is degenerated on the closed curve $\overline{\Gamma}_2$, that is

$$(B^2 - AC)|_{DU\Gamma_1} < 0, \tag{3}$$

$$(B^2 - AC)|_{\Gamma_2} = 0. \tag{4}$$

Let $H_i(x, y) = 0$ be the equation of $\overline{\Gamma}_i$, where $H_i|_D > 0$,

$$H_i \in C^2(\overline{D}),\ H_i|_{\Gamma_i} = 0,\ \nabla H_1|_{\overline{\Gamma}_1} \neq 0,\ \nabla H_2|_{\overline{\Gamma}_2} \neq 0\ \ (i = 1, 2).$$

We suppose that the points P_1 and P_2 are not cusps for ∂D.

Let us formulate the oblique derivative problem [1]. Find the regular solution of the equation (1) $U \in C(\overline{D}) \cap C^1(D \cup \Gamma_1) \cap C^2(D)$ which satisfies the following boundary conditions:

$$\Lambda(U)|_{\Gamma_1} = \left(\frac{\partial U}{\partial l} + dU\right)\Big|_{\Gamma_1} = f_1, \tag{5}$$

$$U|_{\overline{\Gamma}_2} = f_2, \tag{6}$$

where $\frac{\partial U}{\partial l}$ denotes the directional derivative along the vector which makes an acute angle to the internal normal of $\overline{\Gamma}_1$ $d, f_1 \in C(\overline{\Gamma}_1)$, $f_2 \in C(\Gamma_2)$, $d \le 0$ we assume that the components of the unite vector l belong to $C(\overline{\Gamma}_1)$.

We consider the two cases:

$$(AH_{2x}^2 + 2BH_{2x}H_{2y} + CH_{2y}^2)|_{\overline{\Gamma}_2} \neq 0, \tag{7}$$

$$(AH_{2x}^2 + 2BH_{2x}H_{2y} + CH_{2y}^2)|_{\overline{\Gamma}_2} = 0.$$

Lemma. For a solution of the problem (1), (5), (6) of the class $C(\overline{D})\cap C^1(D\cup\Gamma_1)\cap C^2(D)$ the following *apriori* estimate

$$\|U\|_{C(\overline{D})} \leq C^*(\|f_1\|_{C(\Gamma_1)} + \|f_2\|_{C(\overline{\Gamma}_2)} + \|F\|_{C(D)}), \tag{8}$$

holds, where C^* is a positive constant non-dependent on U.

When considering the second case of (7), we will provide that in some neighbourhood of Γ_2 the following representation

$$A H_{2x}^2 + 2BH_{2x}H_{2y} + C H_{2y}^2 = H_2^P G,$$

holds, where $p = \text{const} > 0$, $G > 0$; moreover we assume that one of the next conditions [2]: 1) $0 < p < 1$; 2) $p = 1$, $(1 - IG^1)|_{\overline{\Gamma}_2} > 0$, $I = L(H_2) - cH_2$; 3) $1 < p < 2$, $I|_{\overline{\Gamma}_2} \leq 0$; 4) $p \geq 2$, $I|_{\overline{\Gamma}_2} \leq 0$ holds.

Theorem. *Let the direction l make obtuse angles to the internal normals at the points P_1 and P_2 of $\overline{\Gamma}_2$. Then for any function $f_1 \in C^{1,\alpha}(\overline{\Gamma}_1)$, $f_2 \in C(\overline{\Gamma}_2)$, there exists the unique solution $U(x, y)$ of the problem (1), (5), (6) which belongs to the class $C^{2,\alpha_1}(\overline{D}\setminus\overline{\Gamma}_2)\cap C(\overline{D})$ where $0 < \alpha_1 < \alpha$.*

Note that the uniqueness of solution of the problem (1), (5), (6) for the wider class $C(\overline{D})\cap C^1(D\cup\Gamma_1)\cap C^2(D)$ follows from the *apriori* estimate (8).

I.Javakhishvili Tbilisi State University

REFERENCES

1. *O.A.Oleinik*. Mathem. J. **30**, 3, 1952, 695-702.
2. *M.A.Usanetashvili*. Differential Equations. **21**, 1, 1985, 145-148.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

T.Kadeishvili, S.Saneblidze

On the Iterated Bar Construction

Presented by Corr.Member of the Academy N.Berikashvili, July 12, 1996

ABSTRACT. **For one-connected space X the Adams's bar construction $B(C^*(X))$ describes $H^*(\Omega X)$ just as graded module and gives no information about the multiplicative structure. Thus it is not possible to iterate the bar construction in order to determine cohomology of iterated loop spaces $\Omega^i X$. In this article for n-connected pointed space X a sequence of A(∞) - algebra structures $\{m_i^{(k)}\}$, $k=1, 2,..., n$ is constructed, such that for each $k \leq n$ there exists an isomorphism of graded algebras**

$$H^*(\Omega^k X) \cong (H(B(...(B(B(C^*(X);\{m_i^1\});\{m_i^{(2)}\});...);\{m_i^{(k-1)}\}));m_2^{(k)}*).$$

For 1-connected pointed space X Adams [1] found a natural isomorphism of graded modules

$$H(B(C^*(X)) \cong H^*(\Omega X),$$

where $B(C^*(X))$ is the bar construction of DG - algebra $C^*(X)$. The method can not be extended directly for iterated loop spaces $\Omega^k X$, for $k \geq 2$, since the bar construction $B(A)$ of a DG - algebra A is just a DG - coalgebra; it does not carry a structure of DG - algebra in order to produce the double bar construction $B(B(A))$.

However for $A=C^*(X)$ Baues [2] has constructed an associative product

$$\mu: B(C^*(X)) \otimes B(C^*(X)) \to B(C^*(X)),$$

which turns $B(C^*(X))$ into a DG-algebra and which is geometric: for 1-connected X there exists an isomorphism of graded algebras

$$H(B(C^*(X)) \cong H^*(\Omega X)$$

and for 2-connected X there exists an isomorphism of graded modules

$$H(B(B(C^*(X)) \cong H^*(\Omega^2 X).$$

But again, as it is mentioned in [2], this method can not be also extended for $\Omega^3 X$, since "it is impossible to construct a "nice" product on $B(B(C^*(X))$".

We remark here that in order to produce the bar construction $B(M)$ it is not necessary to have on a DG-module M a strict associative product

$$\mu: M \otimes M \to M;$$

it is sufficient to have a strong homotopy associative product or, what is the same, to have $A(\infty)$-algebra structure on M. This notion was introduced by Stasheff in [12].

$A(\infty)$-algebra $(M, \{m_i\})$ is a graded module M equiped with a sequence of operations

$$\{m_i: M \otimes ...(i\text{–}times)... \otimes M \to M, \quad i=1, 2, 3,...; \quad deg\, m_i = 2-i\},$$

which satisfies the associativity condition

$$\sum_{k=1}^{n-1} \sum_{j=1}^{n-k+1} m_{n-j+1} (a_1 \otimes ... \otimes a_k \otimes m_j (a_{k+1} \otimes ... \otimes a_{k+j}) \otimes ... \otimes a_n) = 0$$

for each $a_i \in M$ and $n>0$.

Such a sequence defines on

$$B(M) = R+s^{-1}M + s^{-1}M \otimes s^{-1}M + ... = \sum_{i=0}^{\infty} \otimes^i s^{-1}M$$

(here R is a ground ring and $s^{-1}M$ is the desuspension of M) the correct differential

$$d_m: B(M) \to B(M)$$

given by

$$d_m(s^{-1}a_1 \otimes ... \otimes s^{-1} a_n) =$$

$$= \sum_{k=1}^{n-1} \sum_{j=1}^{n-k+1} s^{-1} a_1 \otimes ... \otimes s^{-1} a_k \otimes s_{m_j}^{-1} (a_{k+1} \otimes ... \otimes a_{k+j}) \otimes ... \otimes s^{-1} a_n,$$

which is a coderivation with respect to standard coproduct. This DG - coalgebra ($B(M)$; d_m) is denoted as $B(M, \{m_i\})$ and is called the bar construction of $A(\infty)$ - algebra (M, $\{m_i\}$).

In particular an $A(\infty)$ - algebra of type

$$(M, \{m_1, m_2, 0, 0,...\})$$

is just a DG - algebra with differential m_1 and strict associative product m_2 (up to signs). For such an $A(\infty)$ - algebra $B(M, \{m_i\})$ coincides with usual bar construction. For a general $A(\infty)$ - algebra $(M, \{m_i\})$ the first operation m_1: $M \to M$ is a differential, which is a derivation with respect to the second operation m_2: $M \otimes M \to M$; this operation is not necessarily associative, but is homotopy associative (the operation m_3 is the suitable homotopy). Thus it is possible to consider homology of DG - module (M, m_1) and the product m_2 induces on $H(M, m_1)$ the strict associative product m_2^*.

Now we can formulate the main result of this paper.

Theorem 1. *Let X be an n - connected pointed space, then there exists the sequence of* $A(\infty)$ *- algebra structures* $\{m_i^{(k)}\}$, $k=1, 2,..., n$, *such that for each* $k \leq n$ *there exists an isomorphism of graded algebras*

$$H^*(\Omega^k X) \cong$$

$$\cong (H(B(...(B(B(C^*(X);\{m_i^{(1)}\});\{m_i^{(2)}\});...);\{m_i^{(k-1)}\}));m_2^{(k)*}).$$

The proof is based on the following

Theorem 2. *Let* $(M, \{m_i\})$ *be an* $A(\infty)$*-algebra,* (C, d) *be a DG - module and*

$$f: (M, m_1) \to (C, d)$$

be a weak equivalence of DG-modules, assume further, that M and C are connected and free as graded modules. Then there exist

(1) *an* $A(\infty)$ *- algebra structure* $\{m_i'\}$ *on C with* $m_1'=d$;

(2) *a morphism of* $A(\infty)$ *- algebras*

$$\{f_i\}: (M, \{m_i\}) \to (C, \{m_i'\})$$

with $f_1 = f$.

The proof of theorem 2 is based on the Perturbation Lemma [4,5], extended for chain equivalences in [7].

Let us mention some results from literature, dedicated to the problem of iterating of bar construction.

In [9] Khelaia has constructed on $C^*(X)$ the additional structure, which particularly containes the Steenrod's U_1 product, and which is used to introduce in the bar construction $BC^*(X)$ a homotopy associative product, which is geometric: there exists an isomorphism of graded algebras

$$H^*(B(C^*(X)) \cong H^*(\Omega X);$$

it is possible to show, that this product is strong homotopy associative, thus there appears the possibility to produce the next bar construction $B(B(C^*(X)))$. But the additional structure itself is lost in $B(C^*(X))$, thus this structure is not enough to determine a product in $B(B(C^*(X)))$.

Later Smirnov [10], using the techniques of operads, has introduced more powerful additional structure - the E_∞-structure, which can be transferred on $B(C^*(X))$. Thus there appears the possibility of iteration.

The structure of m - algebra introduced by Justin Smith [1] also is transferable on the bar construction and, as it is mentioned in [11], is smaller and has computational advantages against Smirnov's.

It seems that the structure, introduced in this article - the sequence of $A(\infty)$-algebra structures $\{m_i^{(k)}\}$ should be a smallest one, because the $A(\infty)$-algebra structure is a minimal structure necessary to produce the bar construction.

The lack of our structure is that it can not be considered as a sequence of operations from $Hom(\otimes^k C^*(X), C^*(X))$ as in [10] and [11]. In the forthcoming publication we are going to elaborate this lack.

The research described in this publication was made possible in part by Grant N RVC2OO from the International Science Foundation.

A.Razmadze Mathematical Institute
Georgian Academy of Sciences

REFERENCES

1. *J.F.Adams, P.J.Hilton.* Math. Helv. **20**, 1995, 305-330.
2. *H.Baues.* Memoirs of A.M.S. **25**, 230, 1980, 55-106.
3. *E.Brown.* Twisted tensor product. Ann. of Math. (2) **69**, 1959, 223-246.
4. *R.Brown.* The Twisted Eilenberg-Zilber Theorem. Celebrazioni Archimedee del Secolo XX, Simposio di Topologia , 1967, 34-37.
5. *V.K.A.M.Gugenheim.* J. of Math. **3**, 111, 1972, 392-414.
6. *V.K.A.M.Gugenheim. L.Lambe, J.D.Stasheff.* Perturbation Theory in Differential Homological Algebra II. Preprint, 1989.
7. *I.Huebschman, T.Kadeishvili.* Math. Zeitschrift, **207**, 1991, 245-280.
8. *T.Kadeishvili.* Proc. Tbilisi Math. Inst. **107**, 1993, 1-94.
9. *L.Khelaia.* On Some Chain Operations. Proc. Tbilisi Math. Inst. **83**, 1986, 102-115.
10. *V.Smirnov.* Homotopy Theory of Coalgebras. Izv. Akad. Nauk SSSR **49**, 1985, 575-593.
11. *J.Smith.* Iterating the Cobar Construction. Memoirs of A.M.S. **109**, 524, 1994, 1-141.
12. *J.D.Stasheff.* Homotopy Associativity of *H*-spaces. Trans. Amer. Math. Soc. **108**, 1963, 275-312.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

T.Vepkhvadze

General Theta-Functions with Characteristics and Exact Formulae for Binary Quadratic Forms

Presented by Academician G.Chogoshvili, July 18, 1996

ABSTRACT. **The exact formulae for the number of representations of positive integers by the forms $x_1^2 + 19x_2^2$, $4x_1^2 + 2x_1x_2 + 5x_2^2$ and $4x_1^2 - 2x_1x_2 + 5x_2^2$ are obtained.**

Let $A = (a_{ik})$ be an integral square symmetric matrix of s rows and s columns diagonal elements of which are even integers. If x is a variable vector with s components then the matrix product $\frac{1}{2}x'Ax = f(x)$ is a quadratic form. We shall suppose this quadratic form to be definite and positive. The determinant Δ of A is then a positive integer.

An integral vector a with s components will be called special (with respect to A) if Aa is divisible by N, where N is a step of $f(x)$ (it is at least integer for which NA^{-1} is an integral matrix with the even diagonal elements).

Now let τ be a complex variable with a positive imaginary part. Let $P_\nu(x)$ be a spherical function of order ν with respect to A. Let g and h be special vectors with respect to A. Then we define the general theta-function with characteristics [1], by

$$\vartheta_{gh}(\tau, P_\nu(x), f(x)) = \sum_{x \equiv o \pmod g} (-1)^{\frac{h'A(x-g)}{N^2}} P_\nu(x) e^{\frac{\pi i \tau x' A x}{N^2}}. \tag{1}$$

Let

$$\vartheta(\tau; f) = \vartheta_{oo}(\tau, P_\nu(x), f(x)), P_o(x) = 1.$$

Then, according to (1), we have

$$\vartheta(\tau; f) = 1 + \sum_{n=1}^{\infty} r(n; f)\, e^{2\pi i \tau n},$$

where $r(n; f)$ denote the number of representation of positive integer n by f.

The special cases of functions (1) are the Jacobi's theta-functions, simple theta-functions with characteristics and their derivatives. They were used to express the additional members in the exact formulae for the number of representations of integers by some quadratic forms. Using more general functions defined by (1) gives us the opportunity to obtain new formulae and disclose arithmetical meaning of the remainder members in these formulae.

In the paper [2] the modular properties of functions (1) are learned. There exists more general theorem which can be stated as follows:

Theorem 1. *Let $f_1 = f_1(x)$, ..., $f_j = f_j(x)$ be the positive definite quadratic forms in s variables. Let $P_\nu^{(k)}(x)$ be spherical functions with respect to f_k ($k = 1, ..., j$). Let A_k be a matrix of f_k and Δ_k be a determinant of A_k ($k = 1, ..., j$). Let Δ be a determinant of a positive definite quadratic form f in s + 2ν variables. Let N be a step of f.*

Now let $g^{(k)}$ and $h^{(k)}$ be special vectors with respect to $A^{(k)}$ but if $2 \mid \frac{N}{N_K}$ then h_k a vector with even components. Let B_k be arbitrary complex number. Then the function

$$X(\tau) = \sum_{k=1}^{j} B_K \vartheta_{g^{(K)}h^{(K)}}(\tau; P_\nu^{(k)}(x), f_k(x))$$

is an integral modular form of type $\left(-\left(\frac{s}{2}+\nu\right), N, v(L)\right)$, where $v(L)$ is multiplier system with respect to f, if and only if

$$N_K \mid N,\ N_K^2 \mid f_K(g^{(K)}),\ 4N_K \left| \frac{N}{N_K} f_K(h^{(K)}) \right.$$

and for all α and δ that satisfy the condition $\alpha\delta \equiv 1 (\mathrm{mod} N)$, we have

$$\sum_{k=1}^{j} B_k \vartheta_{\alpha g^{(k)}h^{(k)}}(\tau; P_\nu^{(k)}(x), f_k(x))\,(sgn\delta)^\nu \left(\frac{(-1)^{\left[\frac{s}{2}\right]}\Delta_K}{|\delta|}\right) =$$

$$= \left(\frac{(-1)^{\left[\frac{s+2\nu}{2}\right]}\Delta}{|\delta|}\right) \sum_{k=1}^{j} B_k \vartheta_{g^{(k)}h^{(k)}}(\tau; P_\nu^{(k)}(x), f_k(x)).$$

In this paper having used this theorem the exact formulae for the number of representations of positive integers by the forms $x_1^2 + 19x_2^2$, $4x_1^2 + 2x_1x_2 + 5x_2^2$ and $4x_1^2 - 2x_1x_2 + 5x_2^2$ are obtained. The first form was considered by Lomadze [3]. He obtained the formulae with 3 remeinder members. Those were the coefficients of certain cusp forms without arithmetical meaning. The formula which is obtained in this paper has only one remainder member and it's arithmetical meaning is shown. The rest of the two forms are first considered

Theorem 2. *Let* $f_1 = x_1^2 + 19x_2^2$, $f_2 = 4x_1^2 + 2x_1x_2 + 5x_2^2$, $f_3 = 4x_1^2 - 2x_1x_2 + 5x_2^2$, $g = \begin{pmatrix} 0 \\ 38 \end{pmatrix}$, $h = \begin{pmatrix} 38 \\ 0 \end{pmatrix}$. *Then we have*

$$\vartheta(\tau; f_1) = \frac{1}{2}\,\theta(\tau; f_1) + \frac{2}{3}\,\vartheta_{gh}(\tau; P_0(x), f_2),$$

$$\vartheta(\tau; f_2) = \vartheta(\tau; f_3) = \frac{1}{2}\,\theta(\tau; f_1) - \frac{1}{3}\,\vartheta_{gh}(\tau; P_0(x), f_2),$$

where

$$\theta(\tau; f_1) = 2 + \sum_{n=1}^{\infty} \rho(n; f_1)\, e^{2\pi i \tau n}.$$

The function $\rho(n; f)$ may be calculated by the paper [3].

Theorem 3. *Let* $n = 2^{\alpha}19^{\beta}u$, $(u, 38) = 1$. *Let* $r(n; f_k)$ *denote the number of representations of positive integer* n *by the form* f_k $(k = 1,2,3)$. *Then*

$$r(n;f_k) = \frac{1}{3}\left(1+\left(\frac{u}{19}\right)\right)\sum_{d\mid u}\left(\frac{-19}{d}\right) + \nu(n;f_k) \quad \text{if } \alpha = 0;$$

$$= \left(1+\left(\frac{u}{19}\right)\right)\sum_{d\mid u}\left(\frac{-19}{d}\right) \quad \text{if } 2\mid\alpha,\ \alpha > 0;$$

$$= 0 \quad \text{if } 2\nmid\alpha$$

where

$$\nu(n;f_1) = \frac{2}{3}\sum_{n=x_1^2+x_1x_2+5x_2^2}(-1)^{x_1},$$

$$\nu(n;f_2) = \nu(n;f_3) = -\frac{1}{3}\sum_{n=x_1^2+x_1x_2+5x_2^2}(-1)^{x_1}.$$

I.Javakhishvili Tbilisi State University

REFERENCES

1. *T.V.Vepkhvadze*. Bull. Georg. Acad. Sciences, **128**, 2, 1987, 465-468.
2. *T.V.Vepkhvadze*. J. Acta Arithmetica, 53, 1990, 433-451.
3. *G.A.Lomadze*. Izvestia Visshikh Uchebnikh Zavedenii, **11**, 1970, 71-75.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

O.Jokhadze

General Darboux Type Problem for a Third Order Equations with Dominated Lower Terms

Presented by Academician T.Burchuladze, July 23, 1996

ABSTRACT. **General Darboux type problem for a third order equation with dominated lower terms is considered in the case of two independent variables. The Banach space $\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)$, $\alpha\geq 0$, is introduced, where the problem under consideration is investigated. A real number α_0 is found such that the problem is uniquely solvable for $\alpha>\alpha_0$ and is normally solvable in the Hausdorff sense for $\alpha<\alpha_0$. An estimate of the solution of the problem ensuring the solution stability is obtained in the class uniqueness.**

In the plane of independent variables x, y let us consider the third order equation with dominated lower terms

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

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

Straight lines y = const form a double family of characteristics of (1), while x=const a single family.

Let $\gamma_1 : y=\gamma_1(x)$, $0\leq x<+\infty$ and γ_2: $x=\gamma_2(y)$, $0\leq y<+\infty$, -be two simple sufficiently smooth curves, coming out of the origin O(0, 0) of the plane of the variables x, y and located completely at the angle $x\geq 0$, $y\geq 0$. Below we shall assume that $\tau(x)\equiv\gamma_2(\gamma_1(x))$, $x\geq 0$, $\tau(x)<x$, $x>0$ and each of the curves γ_i, $i=1,2$, is either the characteristic of the equation (1) or it has characteristic direction at none of its points. Denote by D the domain lying at the angle $x\geq 0$, $y\geq 0$ and bounded by the curves γ_1, γ_2.

Let P_1^0 and P_2^0 be the points at which γ_1 and γ_2 intersect respectively the characteristics $L_1(P^0)$: $x=x_0$ and $L_2(P^0)$: $y=y_0$, coming out of an arbitrarily taken point $P^0(x_0, y_0)\in D$. The equation (1) will be considered in the rectangular domain $D_0\equiv\{(x, y)$: $0<x<x_0$, $0<y<y_0\}$ bounded by the characteristics $x=0$, $x=x_0$ and $y=0$, $y=y_0$.

For the equation (1) we shall consider a general Darboux type boundary value problem formulated as follows: Find in $\overline{D}_0$ a regular solution u of the equation (1), satisfying on the segments OP_1^0 and OP_2^0 of the curves γ_1 and γ_2 the conditions

$$(M_1u_{xx}+N_1u_{xy}+P_1u_x+Q_1u_y+S_1u)\big|_{OP_1^0}=f_1,$$

$$(M_ku_{xx}+N_ku_{xy}+P_ku_x+Q_ku_y+S_ku)\big|_{OP_2^0}=f_k,\quad k=2,3, \tag{2}$$

where M_i, N_i, P_i, Q_i, S_i, f_i, $i=1,2,3$, - are the given real functions.

It should be noted that the boundary value problem (1), (2) is the natural development of the well-known classical Goursat and Darboux problems (see, e.g.,[1]-[3]) for the second order linear hyperbolic equations with two independent variables on a plane. The initial-boundary and characteristic problems for a wide class of hyperbolic equations of the third and higher order with dominated derivatives are treated in [4]-[8] and in other papers.

Remark 1. The problem (1),(2) has hyperbolic nature because it contains only the derivatives dominated by the $D_x^2 D_y u$

Remark 2. Since y=const is the double family of characteristics for the hyperbolic equation (1), two conditions are given on the segment OP_2^0 of the curve γ_2.

Remark 3. The above mentioned problem can also be formulated for a curvilinear domain bounded by the curves γ_1, γ_2 and the characteristics $L_1(P^0)$, $L_2(P^0)$ of the equation (1) under the same boundary conditions (2). But in that case, as is well known, the solution u of the formulated problem continues into the domain D_0 as the solution of the original problem (1), (2). The problem considered in the doman D_0 is convenient from the standpoint of effective use of the Riemann function.

Let us introduce the following functional spaces

$$\overset{0}{C}_\alpha(\overline{D}_0) \equiv \{v:\ v\in C(\overline{D}_0),\ v(O)=0,\ \sup_{z\neq 0,\, z\in \overline{D}_0} |z|^{-\alpha}|v(z)|<+\infty\},\ z=x+iy,$$

$$\overset{0}{C}_\alpha[0,d] \equiv \{\varphi:\ \varphi\in C[0,d],\ \varphi(0)=0,\ \sup_{0<t\le d} t^{-\alpha}|\varphi(t)|<+\infty\},\ \alpha\ge 0,\ d>0.$$

Obviously, $\overset{0}{C}_\alpha(\overline{D}_0)$ and $\overset{0}{C}_\alpha[0,d]$ will be the Banach spaces with respect to the norms.

$$\|v\|_{\overset{0}{C}_\alpha(\overline{D}_0)} = \sup_{z\neq 0,\, z\in\overline{D}_0} |z|^{-\alpha}|v(z)|,\ \|\varphi\|_{\overset{0}{C}_\alpha[0,d]} = \sup_{0<t\le d} t^{-\alpha}|\varphi(t)|,$$

respectively.

The boundary value problem (1),(2) will be investigated in the space $\overset{0}{C}{}^{2,1}_\alpha(\overline{D}_0) \equiv \{u:\ D_x^i D_y^j u\in \overset{0}{C}_\alpha(\overline{D}_0),\ i=0,1,2,\ j=0,1\}$ which is Banach with respect to the norm $\|u\|_{\overset{0}{C}{}^{2,1}_\alpha(\overline{D}_0)} = \sum_{i=0}^{2}\sum_{j=0}^{1} \|D_x^i D_y^j u\|_{\overset{0}{C}_\alpha(\overline{D}_0)}$.

To consider (1),(2) in the space $\overset{0}{C}{}^{2,1}_\alpha(\overline{D}_0)$ it is required that $a^{ij}\in C^{i,j}(\overline{D}_0)$, $i=0,1,2$, $j=0,1$, $i+j\neq 3$, $f\in \overset{0}{C}_\alpha(\overline{D}_0)$, M_1, N_1, P_1, Q_1, $S_1\in C[0,x_0]$, M_i, N_i, P_i, Q_i, S_i, $\in C[0,y_0]$, $i=2,3$, $f_1\in \overset{0}{C}_\alpha[0,x_0]$, $f_i\in \overset{0}{C}_\alpha[0,y_0]$, $i=2,3$.

Assume that

$$\Delta(y) \equiv \begin{vmatrix} Q_2(y) & N_2(y) \\ Q_3(y) & N_3(y) \end{vmatrix} \neq 0,\ 0\le y\le y_0. \tag{3}$$

and

$$M_1(x)\neq 0,\ 0\le x\le x_0. \tag{4}$$

The case i). The curves γ_1 and γ_2 do not have a common tangent line at the point O and are not the characteristics of the equation (1). Since γ_1 and γ_2 are located completely at the angle $x\ge 0$, $y\ge 0$, we have the inequalities $0<\tau_0<1$, where

$\tau_0 \equiv \tau'(0)$. Set $\sigma \equiv M_1^{-1}(0)\,\Delta^{-1}(0)(M_2N_3Q_1-M_3N_2Q_1+M_3N_1Q_2-M_2N_1Q_3)(0)$ and $\alpha_0 \equiv -\dfrac{\log|\sigma|}{\log \tau_0}$ $(\sigma \neq 0)$, here the subscript – 1 denotes a invertable value.

Theorem 1. Let the conditions (3) and (4) be fulfilled. If $\sigma = 0$, than the problem (1),(2) is uniquely solvable in the space $\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)$ for all $\alpha \geq 0$. If however $\sigma \neq 0$, then the problem (1), (2) is uniquely solvable in the space $\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)$, for $\alpha > \alpha_0$, while for $\alpha < \alpha_0$ the problem (1), (2) is normally solvable in the Hausdorff sense in the space $\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)$ and its index æ $= +\infty$, in particular, the homogeneous problem corresponding to (1), (2) has an infinite number of linearly independent solutions.

The case ii). The curves γ_1 and γ_2 have a common tanget line at the point O, i.e., $\tau_0 = 1$. Then by virtue of the above assumptions the curves γ_1 and γ_2 are not the characteristics of the equation (1). In that case we have the following

Theorem 2. Let the conditions (3) and (4) be fulfilled. If $|\sigma| < 1$, then the problem (1),(2) is uniquely solvable in the space $\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)$ for all $\alpha \geq 0$, while for $|\sigma| > 1$ the problem (1), (2) is normally solvable in the Hausdorff sense in the space $\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)$ for all $\alpha \geq 0$ and its index æ $= +\infty$.

The case iii). Let now at least one of the curves γ_1 or γ_2-be the characteristic of the equation (1) (i.e., $\tau(x) = 0$, $0 \leq x \leq x_0$) Then the following Theorem is valid.

Theorem 3. Let the condition (3) be fulfilled. For the problem (1), (2) to be uniquely solvable in the space $\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)$, $\alpha \geq 0$, for any right-hand sides $f \in \overset{0}{C}_{\alpha}(\overline{D}_0)$, $f_1 \in \overset{0}{C}_{\alpha}[0,x_0]$, $f_i \in \overset{0}{C}_{\alpha}[0,y_0]$, $i = 2, 3$, it is necessary and sufficient that the condition (4) be fulfilled.

It is shown that if the conditions of Theorems 1, 2 and 3 ensuring the unique solvability of the problem (1), (2) are fulfilled, then for the solution u of this problem belonging to the space $\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)$ we have the estimate

$$\|u\|_{\overset{0}{C}{}^{2,1}_{\alpha}(\overline{D}_0)} \leq C\left(\|f_1\|_{\overset{0}{C}_{\alpha}[0,x_0]} + \|f_2\|_{\overset{0}{C}_{\alpha}[0,y_0]} + \|f_3\|_{\overset{0}{C}_{\alpha}[0,y_0]} + \|f\|_{\overset{0}{C}_{\alpha}(\overline{D}_0)}\right),$$

where C-is a positive constant not depending on f, f_i, $i=1, 2, 3$. This estimate immediately implies that the solution of the problem (1), (2) is stable.

The problem of defining the dependence domain for the point $P = P(x,y) \in \overline{D}_0$ is considered. It is shown that for the point $P \in \overline{D}_1 \equiv \{(\xi, \eta): 0 \leq \xi \leq x_0,\ 0 \leq \eta \leq \gamma_1(\xi)\}$ the set

$$D_P^1 \equiv \{(\xi, \eta): 0 \leq \xi \leq x,\ 0 \leq \eta \leq \gamma_1(x)\},$$

is the dependence domain, i.e., the solution u of the problem (1),(2) at the point P is completely defined by the values of the coefficients and the right-hand side of (1) in the closed domain D_P^1, and also by boundary functions values on the segments of the curves γ_1 and γ_2 contained in D_P^1. By a similar reasoning it is shown that if the point P

belongs to $\overline{D}_2 \equiv \{(\xi, \eta): \gamma_2(\eta) \le \xi \le x_0, \gamma_1(\xi) \le \eta \le y_0\}$ or $\overline{D}_3 = \{(\xi, \eta): 0 \le \eta \le y_0, 0 \le \xi \le \gamma_2(\eta)\}$ then the sets $D_P^2 \equiv \{(\xi, \eta): 0 \le \xi \le x, 0 \le \eta \le y\}$ and $D_P^3 \equiv \{(\xi, \eta): 0 \le \xi \le \gamma_2(y), 0 \le \eta \le y\}$ will be the dependence domains of this point, respectively.

A.Razmadze Mathematical Institute
Georgian Academy of Scinces

REFERENCES

1. *G.Darboux*. Thoerie Generale des Surfaces, Paris, 3, 1894.
2. *A.V.Bitsadze*. Some Classes of Partial Equations. Moscow, 1981 448.
3. *S.S.Kharibegashvili*. Dokl. Akad. Nauk SSSR, **280**, 6, 1985, 1313-1316.
4. *D.Colton*. Journal of Differential Equations, **12**, 3, 1972, 559-565.
5. *M.Kh.Shkhanukov*. Differentsial'nie Uravneniya, **18**, 4, 1982, 689-699, (Russian).
6. *S.S.Akhiev*. Dokl. Akad. Nauk SSSR,**271**, 2, 1983, 265-269.
7. *A.P.Soldatov and M.Kh.Shkhanukov*. Dokl. Akad. Nauk SSSR, **297**,3. 1987, 547-552
8. *O.Jokhadze*. Georgian Mathematical Journal, 2, 5, 1995, 469-490.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

M.Khabazi

Weighted Norm Inequalities for Maximal Functions in Orlicz Spaces

Presented by Academician L.Zhizhiashvili, August 22, 1996

ABSTRACT. **The necessary and sufficient condition is derived to fulfill a strong type weighted inequality in Orlicz classes for Hardy-Littlewood maximal function with respect to a nonnegative, continuous Borel measure.**

Let μ be nonnegative, continuous Borel measure on R (i.e. $\mu\{a\}=0$ for every $a\in R$). Given locally integrable function $f\in L^1_{loc}(\mu)$, we define the maximal function $M_\mu f(x)=Mf(x)$ by

$$M_\mu f(x)=\sup_I \frac{1}{\mu I}\int_I |f(x)|d\mu(x),\quad x\in R \tag{1}$$

where the supremum is taken over all intervals I containing x and the ratio is considered zero if $\mu I=0$.

B.Muckenhoupt has proved ([1], theorem 7), the following.

Theorem. *Let w be weight function (a. e. positive, locally integrable function on R) and $1<p<\infty$. Then the following statements are equivalent.*

(i) *There is a constant c, independent of f, such that for every $f\in L^p_w(\mu)$*

$$\int_R |Mf(x)|^p w(x)d\mu(x) \le c\int_R |f(x)|^p w(x)d\mu(x).$$

(ii) *w satisfies $A_p=A_p(\mu)$ condition*:

$$\sup_I\left(\frac{1}{\mu I}\int_I w(x)d\mu(x)\right)\left(\frac{1}{\mu I}\int_I w(x)^{-\frac{1}{p-1}}d\mu(x)\right)^{p-1}<\infty$$

where the supremum is taken over all intervals $I\subset R$.

In this paper similar problems are investigated for $\varphi_w(L)$ classes. By Φ we shall denote the set of all functions $\varphi\colon R^1\to R^1$ which are nonnegative, even and increasing on $(0,\infty)$ such that $\varphi(0+)=0$, $\varphi(\infty)=\infty$. $\varphi_w(L)=\varphi_w(L,\mu)$ is the class of all μ-measurable f functions for which $\int_R \varphi(f(x))\,w(x)dx<\infty$ ([2], [3]). For our purpose we shall need the following definition of quasiconvex functions.

A function ω is called a Young function on $[0,\infty)$ if $\omega(0)=0$, $\omega(\infty)=\infty$, ω is convex and is not identically zero or ∞ on $(0,\infty)$; it may have a jump up to ∞ at some point $t>0$, but in this case it should be left continuous at t.

A function φ is called quasiconvex if there exist a Young function ω and a constant $c>1$ such that

$$\omega(t) \le \varphi(t) \le \omega(c\, t),\ t \ge 0.$$

Now we are ready to formulate our results.

Theorem 1. *Let w be a weight function on R and $\varphi \in \Phi$. Then the following statements are equivalent.*

(i) *φ^{α} is quasiconvex for some $\alpha \in (0,1)$ and $w \in A_{p(\varphi)}$, where*

$$\frac{1}{p(\varphi)} = \inf\{\alpha:\ \varphi^{\alpha} \text{ is quasiconvex}\}.$$

(ii) *There is a constant c, independent of f, such that for every $f \in \varphi_w(L)$*

$$\int_R \varphi\big(Mf(x)\big)w(x)d\mu(x) \le c\int_R \varphi\big(cf(x)\big)w(x)d\mu(x)\,.$$

We must mention that this theorem is analogous of A.Gogatishvili's and V.Kokilashvili's one ([4], theorem 1), proved for homogeneous type spaces. We shall note also that weak-type estimate, obtained by A.Gogatishvili ([5], theorem 5.1) for the same spaces is valid in our case too:

Theorem 2. *Let w be a weight function on R and $\varphi \in \Phi$. Then the following statements are equivalent.*

(i) *There is a constant c_1, independent of f, such that for every $f \in \varphi_w(L)$ and $\lambda > 0$*

$$\varphi(\lambda)w\{x: x \in R,\ Mf(x) > \lambda\} \le c_1 \int_R \varphi\big(c_1 f(x)\big)w(x)d\mu(x)\,.$$

(ii) *φ is quasiconvex and $w \in A_{p(\varphi)}$.*

The following results illustrate the application of theorem 1. Corresponding definitions may be found in [6] and [7].

Theorem 3. *Let $\varphi \in \Phi$, f and w be 2π periodic integrable functions on $[0,2\pi)$ (with respect to Lebesgue measure), $w \ge 0$ and $f(r,\theta)$ denotes the Poisson integral of f $(0 \le r < 1)$. Then the following statements are equivalent.*

(i) *There is a constant c, independent of f, such that*

$$\int_0^{2\pi} \varphi\big(f(r,\theta)\big)w(\theta)d\theta \le c\int_0^{2\pi} \varphi\big(cf(\theta)\big)w(\theta)d\theta,\ 0 \le r < 1.$$

(ii) *φ is quasiconvex and $w \in A_{p(\varphi)}$.*

Theorem 4. *Let $\varphi \in \Phi$, w be a weight function on $[0,\pi)$, $\lambda > 0$, $f(\theta)\sin^{2\lambda}\theta$ is integrable on $[0,\pi)$, $\sum a_n P_n^{\lambda}(\cos\theta)$ is Gegenbauer expansion of f and $f(r,\theta) = $ $= \sum a_n r^n P_n^{\lambda}(\cos\theta)$. Then the following statements are equivalent.*

(i) *There is a constant c, independent of f, such that*

$$\int_0^{\pi} \varphi\big(f(r,\theta)\big)w(\theta)d\theta \le c\int_0^{\pi} \varphi\big(f(\theta)\big)w(\theta)d\theta,\ 0 \le r < 1.$$

(ii) *φ is quasiconvex and there exists $k>0$ such that for every $I \subset [0,\pi)$*

$$\left(\int_I w(\theta)d\theta\right)\left(\int_I (w(\theta))^{-\frac{1}{p(\varphi)-1}}(\sin\theta)^{2\lambda p'(\varphi)}d\theta\right)^{p(\varphi)-1} \le k\left(\int_I \sin^{2\lambda}\theta d\theta\right)^{p(\varphi)},$$

when $p(\varphi) > 1$ and

$$\int_I w(\theta)d\theta \le k\left(\int_I \sin^{2\lambda}\theta d\theta\right)\operatorname*{essinf}_{y\in I}\left(w(y)\sin^{-2\lambda}y\right),$$

when $p(\varphi)=1$.

Finally we shall note that theorem 3 is valid for Cezaro means as well.

A. Razmadze Mathematical Institute
Georgian Academy of Sciences

REFERENCES

1. *B.Muckenhoupt.* Trans. Amer. Math. Soc. **165**, 1973, 207-226.
2. *M.A.Krasnoselski, Ya.B.Ruticki.* Convex functions and Orlicz spaces. Noordhoff (Groningen, 1961).
3. *V.Kokilashvili, M.Krbec.* Weighted inequalities in Lorentz and Orlicz spaces. World Scientific, 1991.
4. *A.Gogatishvili, V.Kokilashvili.* Proc. Georg. Acad. Sci. math. **1**, 6, 1993, 721-755.
5. *A.Gogatishvili.* Proc. Razm. Math. Inst. **101**, 1992. 72-79.
6. *A.Zygmund.* Trigonometrical series. Cambridge Univ. Press, New York, 1959.
7. *B.Muckenhoupt, I.Stain.* Trans. Amer. Math. Soc. **118**, 1965, 17-92.

ᲡᲐᲥᲐᲠᲗᲕᲔᲚᲝᲡ ᲛᲔᲪᲜᲘᲔᲠᲔᲑᲐᲗᲐ ᲐᲙᲐᲓᲔᲛᲘᲘᲡ ᲛᲝᲒᲔᲑᲐ, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MATHEMATICS

S.Saneblidze

The Formula Determining an A_∞-Coalgebra Structure on a Free Algebra

Presented by Corr. Member of the Academy N.Berikashvili, July 25, 1996.

ABSTRACT. **A formula is obtained for an A_∞-coalgebra structure in a free algebra which is defined by its restriction to multiplicative generators.**

Let (A, d) be a differential graded algebra (dga) with a non - associative coproduct $\nu: A \to A \otimes A$. Let ν extend to an A_∞ - coalgebra structure in the sense of Stasheff [1]: $\nu = \nu^0, \nu^1, \dots, \nu^n, \dots,$

$$d\nu^n + \nu^n d = \sum_{p+q=n+1,\ 0\le i\le q+1} \nu_i^p \nu_0^q,$$

where

$$\nu_i^n = 1^{\otimes i} \otimes \nu^n \otimes 1^{\otimes m-i-1} : A^{\otimes m} \to A^{\otimes m+n+1}$$

For $m>i$.

Here we want to consider the cases when each ν^n is determined by its restriction to multiplicative generators of A. It is easy to say what ν^0 and ν^1 might be: one can require that ν^0 is a dga map and ν^1 is a derivation homotopy between $(\nu^0 \otimes 1)\nu^0$ and $(1 \otimes \nu^0)\nu^0$, i.e.

$$\nu^1(ab) = \nu^1(a)(\nu^0 \otimes 1)\, \nu^0(b) + (1 \otimes \nu^0)\nu^0(a)\nu^1(b), \quad a,b\in A.$$

But difficulties arise by defining higher homotopies ν^n, $n \ge 2$.

Below we establish the formulae for this and give the cases when such structures canonically appear. First for a composition $\nu_{i_1}^{n_1} \dots \nu_{i_{m+1}}^{n_{m+1}}$ with $i_1 > i_2 > \dots > i_{m+1}$ denote $l_q = i_{q-1} - i_q - 1$, $q > 1$, and $l_1 = n_1 + \dots + n_{m+1} - l_2 - \dots - l_{m+1}$. Now define

$$\nu^k(ab) = \sum_{k=n+m} \nu_{i_1}^{n_1} \dots \nu_{i_{m+1}}^{n_{m+1}}(a)\, \nu_{j_1}^{m_1} \dots \nu_{j_{n+1}}^{m_{n+1}}(b)$$

where $n = n_1 + \dots + n_{m+1}$, $m = m_1 + \dots + m_{n+1}$, $i_1 > i_2 > \dots > i_{m+1}$,

$$t(j) - \lambda(j) \le m_j \le s_j - \lambda(j),$$
$$s_j = m - 1 - (m_1 + \dots + m_{j-1}).$$

$\lambda(j)$ is integer p, $1 \le p \le m+1$, such that

$$l_{m+1} + \dots + l_{p+1} < j \le l_{m+1} + \dots + l_{p+1} + l_p,$$
$$t(j) = \inf\{q \mid \lambda(j) \le q \le s_j,\ n_q^j \ne 0\},$$

where $n_i^1 = n_i$ and

$$n_i^{j+1} = \begin{cases} n_i^j & \text{if } i < \lambda(j), \\ n_{\lambda(j)}^j + n_{\lambda(j)+1}^j + \dots + n_{\lambda(j)+m_j}^j - 1 & \text{if } i = \lambda(j), \\ n_{i+m_j}^j & \text{if } i > \lambda(j), \end{cases}$$

$1 \le i \le s_{j+1}$.

$$j_r = s_{r+1} - e_r - 1,$$

where $e_r \ge 0$ is the number of i_q's, $1 \le q \le m$ with $i_q - q \ge r$, $1 \le r \le n$.

For example

$$\nu^2(ab)=\nu^2(a)\,\nu_0^0\;\nu_0^0\;\nu_0^0(b)+\nu_1^1\;\nu_0^0(a)\,\nu_0^1\;\nu_0^0(b)+\nu_1^1\;\nu_0^0(a)\,\nu_1^0\;\nu_0^1(b)+$$

$$+\nu_1^0\;\nu_0^1(a)\,\nu_0^1\;\nu_0^0(b)+\nu_2^0\;\nu_0^1(a)\,\nu_0^0\;\nu_0^1(b)+\nu_2^0\;\nu_1^0\;\nu_0^0(a)\,\nu^2(b).$$

Thus ν^2 is not a derivation. However it is well defined by its derivation part (as well as other higher homotopies).

We will refer to a dga A with such A_∞-coalgebra structure as *an A_∞ - Hopf algebra.*

Theorem. *Let A be a free dga such that Hom $(A, A^{\otimes n})$ is acyclic for each $n>2$. Then any dga map $A \to A \otimes A$ inducing an associative coproduct on the homology on A extends to an A_∞ - Hopf algebra structure on A.*

There are some important cases of A with an A_∞ - Hopf algebra structure.

Case 1. A is the cobar-construction $\Omega C_*(X)$ of the simplicial singular complex $C_*(X)$ of a topological space X. The A_∞ - coalgebra structure describes the coalgebra structure of the homology of the loop space ΩX . In fact, there exists an associative coproduct on $\Omega C_*(X)$ ([2]).

Case 2. A is the cobar-construction $\Omega C'_*(X)$ of the cubical singular complex $C'_*(X)$ of a topological space X. The A_∞ - coalgebra structure describes the coalgebra structure of the homology of the loop space ΩX.

Case 3. A is a bigraded algebra resolution RH of a graded Hopf algebra H [3], and the A_∞ - coalgebra structure is compatible with the resolution map $RH \to H$.

The research described in this publication was made possible in part by Grant N RVC200 from the International Science Foundation

A.Razmadze Mathematical Institute
Georgian Academy of Sciences

REFERENCES

1. *T.Kadeishvili.* Russian Math. Survey, **35**, 1980, 231-238.
2. *H.Baues.* Memoirs AMS, 230, 1980.
3. *S.Saneblidze.* Proc. A.Razmadze Math. Inst., **111**, 1994.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

A.Meskhi

Two-Weight Inequalities for Potentials on Homogeneous Groups

Presented by Academician L.Zhizhiashvili, March 6, 1996

ABSTRACT. **In some special case some necessary and sufficient conditions are found for a pair of weight functions, ensuring the validity of two - weight inequalities for potentials defined on homogeneous groups.**

Despite the fact that two-weight problems for potential have already been solved [1-4], it is important to seek new alternative criteria which are simpler for verification and more convenient for application even in those cases when some solutions have already been obtained.

In this paper we consider two-weight inequalities of strong type for operator of potential type T_α defined on homogeneous groups. In some special case we have found necessary and sufficient conditions for boundedness of operator T_α between two-weight L^p spaces. The similar problems were considered in [5], [6], but in this paper more general classes of weight functions are given ensuring validity of two-weight inequalities for operator T_α.

A homogeneous group is a connected and simply connected nilpotent Lie group G whose Lie algebra g is endowed with a family of dilation $\delta_t = \exp(A \ln t)$, where $t > 0$ and A is an operator from g to g of type

$$\begin{bmatrix} \lambda_1 & 0 & 0 & \dots & 0 \\ 0 & \lambda_2 & 0 & \dots & 0 \\ & & \vdots & & \\ 0 & 0 & \dots & 0 & \lambda_k \end{bmatrix} \quad \lambda_i > 0,\ i = 1, 2, \dots, k.$$

The maps $\exp \circ \delta_t \circ \exp^{-1}$ are group of automorphisms of G: we shall denote them also by δ_t and call them delations on G.

The number $Q = trA$ is called a homogeneous dimension of G.

Euclidean spaces and Heisenberg group H^n are examples of Homogeneous groups ([7], [8]).

We define a homogeneous norm on G to be a continuous function $x \to r(x)$ from G to $[0, \infty)$ which is C^∞ on $G \setminus \{e\}$ and satisfies the following conditions:

1. $r(x) = 0 \Leftrightarrow x = e$,
2. $r(x) = r(x^{-1})$ $\forall x \in G$,
3. $r(\delta_t x) = tr(x)$ $\forall x \in G$,
4. There is a constant C_0 such that for every $x, y \in G$

$$r(xy) \le C_0(r(x) + r(y)).$$

If $x \in G$ and $t > 0$ we define

$$B(x, t) = \{y \in G: r(xy^{-1}) < t\},$$

and we call $B(x, t)$ the ball of radius t about x.

We fix the normalization of Haar Measure on G by requiring that the measure of $B(e, 1)$ be 1. We shall denote the measure of any measurable set $E \subset G$ by $|E|$, and the integral of a function $f\colon G \to R^1$ with respect to Haar measure by $\int_G f(x)dx$.

A locally integrable function $w\colon G \to R^1$, which is positive almost everywhere, is called a weight function. Denote by $L^p_w(G)$ a set of measurable functions $f\colon G \to R^1$, with finite norm

$$\|f\|_{L^p_w(G)} = \left(\int_G |f(x)|^p w(x)dx\right)^{\frac{1}{p}}, \; 1 \le p < \infty.$$

Definition 1. We say that the weight $w \in A_p(G)$, $1 < p < \infty$, if there is a constant C_p, such that for every ball $B \subset G$

$$\frac{1}{|B|}\int_B w(x)dx\left(\frac{1}{|B|}\int_B w^{1-p'}(x)dx\right)^{p-1} < C_p, \; p' = \frac{p}{p-1}.$$

Definition 2. The function $\rho\colon G \to R^1_+$ is called the radial function, if there is a function $\beta\colon R^1_+ \to R^1_+$ such that $\rho(x) = \beta(r(x))$ (we shall use the symbol ρ instead of β).

For function $f\colon G \to R$ define the potential

$$T_\alpha f(x) = \int_G \frac{f(y)}{(r(xy^{-1}))^{Q-\alpha}}dy, \; 0 < \alpha < Q.$$

The one-weight problems for operator T_α were solved in [9].

Theorem 1. *Let* $0 < \alpha < Q$, $1 < p < \frac{Q}{\alpha}$, $\frac{1}{p} - \frac{1}{q} = \frac{\alpha}{Q}$, *the radial function* $\rho \in A_{1+\frac{q}{p'}}(G)$, *$\sigma$ and u be positive increasing functions on* $(0, \infty)$. *We put* $v = \sigma\rho$, $w = u\rho$.

If the pair (v, w) *satisfies the condition*

$$\sup_{t>0}\left(\int_t^\infty v(\tau)\tau^{-1-\frac{Qq}{p'}}d\tau\right)^{\frac{p}{q}}\left(\int_0^{\frac{t}{2}}\left(\rho^{-\frac{\alpha p}{Q}}(\tau)w(\tau)\right)^{1-p'}\tau^{Q-1}d\tau\right)^{p-1} < \infty, \tag{1}$$

then there exists a constant $C_1 > 0$ such that the inequality

$$\left(\int_G \left|T_\alpha\left(f \cdot \rho^{\frac{\alpha}{Q}}\right)(x)\right|^q v(r(x))dx\right)^{\frac{1}{q}} \le C_1\left(\int_G |f(x)|^p w(r(x))dx\right)^{\frac{1}{p}} \tag{2}$$

holds for every $f \in L^p_w(G)$.

Theorem 2. *Let* $0<\alpha<Q$, $1<p<\frac{Q}{\alpha}$, $\frac{1}{p}-\frac{1}{q}=\frac{\alpha}{Q}$. σ *and* u *be positive decreasing functions on* $(0,\infty)$, *the radial function* $\rho\in A_{1+\frac{q}{p'}}(G)$. *We put* $v=\sigma\rho$, $w=u\rho$. *If the pair* (v,w) *satisfies the condition*

$$\sup_{t>0}\left(\int_0^{\frac{t}{2}} v(\tau)\tau^{Q-1}d\tau\right)^{\frac{p}{q}}\left(\int_t^{\infty}\left(\rho^{-\frac{\alpha p}{Q}}(\tau)w(\tau)\right)^{1-p'}\tau^{-1-\frac{Qp'}{q}}d\tau\right)^{p-1}<\infty, \qquad (3)$$

then there exists a constant $C_2>0$ *such that the inequality (2) (for constant* C_2*) holds for every* $f\in L^p_w(G)$.

Theorem 3. *Let* $0<\alpha<Q$, $1<p<\frac{Q}{\alpha}$, $\frac{1}{p}-\frac{1}{q}=\frac{\alpha}{Q}$. *If for the radial functions* ρ, v *and* w *the inequality* (2) *holds, then they satisfy conditions* (1), (3).

A.Razmadze Mathematical Institute
Georgian Academy of Sciences

REFERENCES

1. *V.Kokilashvili, M.Gabidzashvili.* Preprint 45. Math. Inst. Czech. Acad. Sci., Prague, 1989.
2. *E.T.Sawyer, R.L.Wheeden.* Amer. J. Math. **114**, 1992, 813-84.
3. *A.Gogatishvili, V.Kokilashvili.* In: Topological Vector Spaces, Algebras and Related Areas. Longman. Pitman Research Notes in Mathematics, **316**, 1994, 251-362.
4. *I.Genebashvili, A.Gogatishvili, V.Kokilashvili.* Georgian Math. J. **1**, 1, 1994, 9-29.
5. *S.K.Abdulaev.* Dokl. Akad. Nauk SSSR, **283**, 4, 1983, 777-780.
6. *V.S.Guliev.* Dokl. Akad. Nauk SSSR, **304**, 6, 1989, 1289-1293.
7. *G.B.Folland, E.M.Stein.* Hardy Spaces on Homogeneous Groups, Princeton, New Jersey, 1982.
8. *E.N.Taylor.* Noncommutative Harmonic Analysis. Amer. Math. Soc. Providence, 1986.
9. *B.Muckenhoupt, R.Wheeden.* Trans. Amer. Math. Soc. **192**, 1974, 261-274.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

CYBERNETICS

M.Akhobadze, D.Zangurashvili

Optimization Problem for Tare Turnover

Presented by Academician M.Salukvadze, August 8, 1996

ABSTRACT. **Mathematical model of the tare turnover is described in the article. The algorithm for the maximal profit is obtained.**

As it is known the tare turnover can make some profit to his owner. It happens when the price of the tare is greater than a consumer tare purchase price. In such case each turnover of the tare unit makes the profit to his owner, which is equal to the difference of the sale and the purchase prices. When the sale price is fixed, the less is the purchase price, the more is the profit. On the other hand, the purchase price has not to be taken too low, because in such case the turnover of the tare will be low, that in its turn will affect the profit. Hence, it is an interesting problem to determine such tare purchase price for which the profit obtained by the tare turnover will be maximal.

Consider the system consisting of three objects, which will be called as the tare supplier, the firm and the consumer in this article. The firm makes the production and it is necessary to have the tare to reach the consumer. The sheme of the tare circulation is shown in the Fig.1.

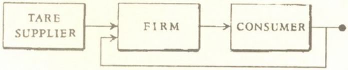


Fig. 1.

Suppose we are intresting in the maximization of the profit during the period of n time units. Let x_1 be the quantity of the tare at the firm and x_2 be the tare at the consumer's disposal in the beginning of the considered time period. Let k_i be the quantity of the tare received by the firm from the tare supplier by the price p during the i-th time interval. Besides α suppose percentages of the available tare is broken in the firm-consumer chain during this time interval. Let the firm sell the tare to the consumer by price u and let x be the quantity of the tare sold in this price in a unit time interval. Suppose that the quantity $f(v)$ of the tare returning to the firm by the consumer in a unit time interval is demanded in the purchase price as follows:

$$f(v) = av^2 + b$$

where a and b are some positive numbers.

Assume that there are z_1 quantity of the tare at the firm's and z_2 – at the consumer's disposal in the beginning of the i -th time interval. Let purchase price v be established in this time interval. There will be

$$\Phi_i(z_1, z_2, v) = \left(1 - \frac{\alpha}{100}\right)[z_1 - \min(z_1, x) + \min(z_2, av^2+b)] + k_i \qquad (1)$$

quantity of the tare and

$$W_i(z_1, z_2, v) = u\min(z_1; x) - v\min(z_2; av^2+b) - k_ip \qquad (2)$$

profit at the firm's disposal at the end of this time interval.

Each i and z_1 uniquely determine z_2. Indeed, the tare quantity in the firm-consumer chain at the end of the i-th time interval is given by the formulae

$$s_0 = x_1 + x_2, \quad s_i = \left(1 - \frac{\alpha}{100}\right) s_{i-1} + k_i, \quad 1 \le i \le n \tag{3}$$

Therefore it can be defined

$$\Phi_i(z_1, v) = \Phi_i(z_1, s_{i-1} - z_1, v) \tag{4}$$

$$W_i(z_1, v) = W_i(z_1, s_{i-1} - z_1, v) \tag{5}$$

Fixed n, x_1, x_2, α, u, x, a, k_1, $k_2,\ldots, k_n$, v_1, $v_2,\ldots, v_n$ parameters, where v_i is the purchase price during the i-th time interval, uniquely determine the quantity y_i of the tare in the firm at the end of the i-th interval ($i = 1, 2,\ldots, n$) and also determine the firm's profit during the period of n time interval:

$$y_0 = x_1, \; y_i = \Phi_i(y_{i-1}, v_i) \quad (i = 1, 2,\ldots, n) \tag{6}$$

$$W = \sum_{i=1}^{n} W_i\,(y_{i-1}, v_i). \tag{7}$$

Notice that it is not all the same for the firm what the quantity of the tare will be there at the end of $[t_0; t_0 + n]$ time interval. So we must take into account the so-called "economic horizon" condition:

$$y_n \ge x_3 \tag{8}$$

where x_3 is some fixed positive number. Besides let's limit prices $v_1, v_2,\ldots, v_n$:

$$u_1 \le v_1, v_2,\ldots, v_n \le u_2 \tag{9}$$

Thus we obtain W's maximization problem with (8)-(9) conditions.

Notice that it is sufficiently to be limited only by discrete values of the tare quantity and price for the firm's needs. Then the tare quantity at the firm's disposal at the i-th time interval end can take only finite quantity of values. If it is assumed h for the tare quantity step, then these values are

$$0, h, 2h,\ldots, l_i h \tag{10}$$

where $l_i = D(s_i, h)$

$$D(x, y) = \begin{cases} \left[\dfrac{x}{y}\right], & \text{if} \quad x - y\left[\dfrac{x}{y}\right] \le \dfrac{y}{2} \\[2ex] \left[\dfrac{x}{y}\right] + 1, & \text{if} \quad x - y\left[\dfrac{x}{y}\right] > \dfrac{y}{2} \end{cases}$$

If it is assumed d for the price step, then the possible values for $v_1, v_2,\ldots,v_n$ are

$$u_1,\; u_1 + d,\; u_1 + 2d,\; \ldots,\; u_1 + cd \tag{11}$$

where $c = \left[\dfrac{u_2 - u_1}{d}\right]$. (1)-(5) formulae take the form:

$$\Phi_i(m, r) = D\,[(1 - \frac{\alpha}{100})(mh - \min(mh; x) + \min(s_{i-1} - mh; a(u_1 + rd)^2 + b) + k_i); h] \tag{12}$$

$$W_i(m, r) = u\min(mh; x) - (u_1 + rd)\min(s_{i-1} - mh; a(u_1 + rd)^2 + b) - k_i p \tag{13}$$

for each m and r, where $0 \le m \le l_i$, $0 \le r \le c$. (6)-(7) formulae change:

$$m_1 = D\,[(1 - \frac{\alpha}{100})(x_1 - \min(x_1; x) + \min(x_2; a(u_1 + r_1 d)^2 + b) + k_1; h] \tag{14}$$

$$m_i = \Phi_i(m_{i-1}, r_i) \quad (i = 2, 3, \ldots, n) \tag{15}$$

$$W_1 = u \min(x_1; x) - (u_1 + r_1 d) \min(x_2; a(u_1 + r_1 d)^2 + b) - k_1 p) \tag{16}$$

$$W = W_1 + \sum_{i=2}^{n} W_i(m_{i-1}, r_i) \tag{17}$$

(8) and (9) conditions take the form:

$$m_n h \geq x_3, \tag{18}$$

$$0 \leq r_1, r_2, \ldots, r_n \leq c \tag{19}$$

and finally we obtain W's (it is given by (12)-(17) maximization problem with (18)-(19) conditions.

Solution of the obtained problem by dynamic programming method [1] gives the following

Algorithm. Find $T(j, k)$ for each (j, k) $(0 \leq j \leq n,\ 0 \leq m \leq l_{i-1}$ and $R(i, m)$, $G(j, k)$ for some (i, m) $(1 \leq i \leq n,\ 0 \leq m \leq l_{i-1})$ and some (j, k) by the formulae

$$G(n, k) = 0,\ T(n, k) = \begin{cases} 0, & if \quad kh < x_3 \\ 1, & if \quad kh \geq x_3. \end{cases}$$

If for each r $(0 \leq r \leq c)$ $T(j+1, \Phi_{j+1}(k, r)) = 0$, then $T(j, k) = 0$, in the other case $T(j, k) = 1$, and determine

$$G(j, k) = \max_{\substack{0 \leq r \leq c \\ T(j+1, \Phi_{j+1}(k, r)) = 1}} (G(j+1, \Phi_{j+1}(k, r)) + W_{j+1}(k, r)$$

$$R(i, m) = arg \max_{\substack{0 \leq r \leq c \\ T(i, \Phi_i(m, r)) = 1}} (G(i, \Phi_i(m, r)) + W_i(m, r)).$$

If it turns out that $T(0, 0) = 0$, then we draw a conclusion, that (18) condition is not fulfilled for each purchase prices. If $T(0, 0) = 1$, then the optimal purchase prices $v_i (1 \leq i \leq n)$ are given by the formulae:

$$v_1 = u_1 + R(1,0)d,\ m_1 = \Phi_1(0, R(1,0))$$

$$v_{i+1} = u_1 + R(i+1, m_i)d,\ m_{i+1} = \Phi_{i+1}(m_i, R(i+1, m_i)).$$

For these prices the tare quantity at the firm's disposal will be

$$x_i = m_i h$$

and obtained profit will be

$$W_i = W_i(m_{i-1}, R(_i, m_{i-1})).$$

The whole maximal profit during the considered period of time unit will be $W = G(0,0)$, which is clearly equal to

$$W = \sum_{i=1}^{n} W_i .$$

(It is generally possible that W is negative showing the inevitable loss and this minimal loss will be reached when the prices are v_i)

Georgian Technical University

REFERENCES

1. *M.Minu.* Matematicheskoie Programmirovanie, M., 1990.

ᲡᲐᲥᲐᲠᲗᲕᲔᲚᲝᲡ ᲛᲔᲪᲜᲘᲔᲠᲔᲑᲐᲗᲐ ᲐᲙᲐᲓᲔᲛᲘᲘᲡ ᲛᲝᲛᲐᲛᲔ, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

PHYSICS

G.Jandieri, Z.Sumnanidze, Zh.Diasamidze

To the Theory of Molecule Motion in a Slowing-Down System

Presented by Academician G.Kharadze, November 21,1996

ABSTRACT. **The dynamic of deceleration of a symmetrical top molecule motion in a slowing down annular parabolic system is studied in this paper. Analytical and numerical investigations of parametrical interaction between periodical external electric field and ammonia molecule are considered in detail.**

It is well known that some effects give contribution to a line width of single molecule of a gaseous media. Inhomogeneous broadening of spectral lines occuring by the Doppler effect is connected with particle motion. The method of substantial variation of molecule's emission and absorption lines shape was suggested in [1]. The idea is based on a capturing of a slowly moving molecules in a standing light wave which is not in resonance with any quantum transitions of molecule and plays only the role of trebly periodic potential field. Molecule motion in a plane standing light wave is described by the oscillator equation.

Mechanism of deceleration of a symmetrical top molecule having big electrical dipole moment is considered in this paper. The deceleration of ammonia molecule is carried out by means of parametrical interaction with an external electric field created by circular parabolic slowing down system [2]. The extraction of energy from a molecule by means of temporal varying external electric field is the physical meaning of deceleration method.

Equation of molecule motion

It is well known that availability of a constant electrical dipole moment d of a symmetrical top pattern molecule gives rise to the Stark effect on a rotational levels in a homogeneous external elecric field. The energy of such molecule in a constant applied electric field is [3]:

$$W = W_0 - d\frac{KM_J}{J(J+1)}E, \quad (K \neq 0) \tag{1}$$

where J, M_J, K are rotational quantum numbers; M_J and K are the projections of an angular momentum on the direction of an external field and on a molecule axis, respectively; E is an electric field strength. The interaction of an external electric field and the molecule occur by means of "effective dipole moment":

$$d_{\text{eff}} = d\frac{KM_J}{J(J+1)} \tag{2}$$

On the other hand the energy of a slightly symmetrical molecule in an external electric field may be written as:

$$W = W_0 \pm \left[\left(\frac{Q}{2}\right)^2 + \left(Ed_{eff}\right)^2\right]^{1/2} \tag{3}$$

where $Q = W_1^0 - W_2^0 = h\nu_0$ is an energy difference between nonexcited energy states; ν_0 is the radiation frequency of molecule; the signs (±) refer to the molecule on an upper and lower energetic levels, respectively.

Equation of motion of molecule having mass m may be written in the following form

$$\ddot{z} = \pm\frac{1}{m}Ed_{eff}^2\left[\left(\frac{Q}{2}\right)^2 + \left(Ed_{eff}\right)^2\right]^{-1/2}\frac{\partial E}{\partial z} \tag{4}$$

Knowledge of a spatial structure of the external field allows us to investigate the character of molecule motion in the external inhomogeneous fields. Further we'll consider the motion of molecule placed on upper energetic level.

Molecule motion-dynamic in a parabolic slowing-down system

Let molecule moves through a series of equidistant thin annular rings with alternate in sign electrostatic potentials [4]. Potential distribution inside the annular system may be approximate as [4]

$$U(r,z) = U_0 + U(r)\cos kz \tag{5}$$

where $U(r) = U_f\dfrac{I_0\left(\frac{2\pi}{l}r\right)\sin\frac{\pi}{2}\left(l - \frac{2\delta}{l}\right)}{I_0\left(\frac{2\pi}{l}R\right)\frac{\pi}{2}\left(l - \frac{2\delta}{l}\right)}$, U_f is the focusing potential i.e. the potential of focusing electrodes with respect to the mean acceleration potential U_0; R and δ are radius and thickness of the rings, respectively; $K = 2\pi/l$; l is a spatial period; I_0 is the modified Bessel function of the zeroth order of the first kind. The potential $U(r,z)$ of an axial symmetric electric field is a function only two coordinates and does not depend on a polar angle φ. Longitudinal component of an axially-symmetrical electic field strength is

$$E_z = -\frac{\partial U(r,z)}{\partial z} = 2\pi\frac{U_f}{l}\frac{I_0\left(\frac{2\pi}{l}r\right)\sin\frac{\pi}{2}\left(1 - \frac{2\delta}{l}\right)}{I_0\left(\frac{2\pi}{l}R\right)\frac{\pi}{2}\left(1 - \frac{2\delta}{l}\right)}\sin kz. \tag{6}$$

Let a thickness of the rings be very small ($\delta = 0$) and spatial period of the system be much greater than the ring radius ($l >> R$).

If the ring radius depends on a coordinate z, in a paraxial approximation the expression for an electric field strength may be written as:

$$E(z) = 4\frac{U_f}{l}\left[1 + \left(\frac{\pi}{l}\right)^2 R^2(z)\right]^{-1}\cos\left(2\pi M\frac{L}{l}\right), \tag{7}$$

where $R(z) = R_0 - r_1 M^2$, r_0 and $R_0 = r_0 + r_1$ are the least and the greatest radii of the rings; $R_0 < l$, $2L$ is the length of the whole system, $M = z/L < 1$.

$$\ddot{z} = -\nabla_z \Pi(z) f^2(t) \tag{8}$$

where $\Pi_1(z) = \frac{1}{Qm} d_{eff}^2 E^2$, $\Pi_2(z) = \frac{1}{m} d_{eff} E$ for weak (index 1) and strong (index 2) electric fields. Potential $\Pi(z)$ is proportional to the square of electric field strength in the case of weak external field and it linearly depends on E in the case of strong field. The function $\Pi(z)$ has the form as shown in Fig. 1. Temporal dependence expressed in terms of the function $f^2(t)$ makes this system parametrically. Parametrical exchange of energy between the molecule and an external field causes energy extraction from molecule which leads to its deceleration.

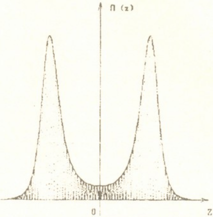


Fig. 1

Substituting equation (3) into equation (8), neglected by system periodicity ($\cos(2\pi z/l) = 1$) and taking into account that parameter M is small, we obtain parametric equation of molecule motion inside the parabolic annular system in the case of weak and strong external electric fields:

$$\ddot{z} + \Omega_{1,2}^2 f^2(t) z = 0 \qquad (9)$$

$$\Omega_1^2 = \frac{2}{Qm}\left(\frac{8\pi U_0 d_{eff}}{l^2 L}\right)^2 R_0 r_1 \left[1 + \left(\frac{\pi}{l}\right)^2 R_0^2\right]^{-3}$$

$$\Omega_2^2 = \left(\frac{4\pi}{lL}\right)^2 \frac{U_0 d_{eff}}{ml} R_0 r_1 \left[1 + \left(\frac{\pi}{l}\right)^2 R_0^2\right]^{-2}$$

$\Omega_{1,2}$- is the own frequency of the system. Time average equation like equation (9) was obtained in [5]. If $f(t) = 1$, the molecule oscillates harmonically inside the annular system with $\Omega_{1,2}$ frequency independently on its initial velocity.

Let the potential on rings exponentially decays in time as $f^2(t)=\exp(-2\varepsilon t)$ ($\varepsilon/\Omega < 1$), $T_0 = 1/\varepsilon$ is an external temporal scale. Dropping the indices (9) reduces to

$$\ddot{z} + \Omega^2 \exp(-2\varepsilon t) z = 0 \qquad (10)$$

general solution of which is

$$z(t) = C_1 J_0(\xi e^{-\varepsilon t}) + C_2 N_0(\xi e^{-\varepsilon t}) \qquad (11)$$

where $J_0(x)$ and $N_0(x)$ are Bessel functions of the first and second kinds of the zeroth order; $\xi = \Omega/\varepsilon$, C_1 and C_2 arbitrary constants must be determined applying the initial conditions: at $t = 0$, $z = 0$ and $\dot{z} = V_0$.

Using asymptotic representation of the functions $J_0(x)$ and $N_0(x)$ for the large argument and taking into account that parameter $\varepsilon/\Omega < 1$ is small, we obtain

$$z(t) = V_0\left(\frac{\pi}{2\Omega\varepsilon}\right)^{1/2}\left[\sin\left(\xi-\frac{\pi}{4}\right)J_0\left(\xi e^{-\varepsilon t}\right)-\cos\left(\xi-\frac{\pi}{4}\right)N_0\left(\xi e^{-\varepsilon t}\right)\right] \tag{12}$$

This solution is exact and holds uniformly for all $t \geq 0$.
At small time intervals when the condition

$$e^{\varepsilon t}/\xi << 1$$

is fulfilled, the solution (12) may be written in the form

$$z(t)= \frac{V_0}{\Omega}e^{\varepsilon t/2}\sin\left[\xi\left(1-e^{-\varepsilon t}\right)\right] \tag{13}$$

It is evident that the process is oscillatory but not periodical.

Mathematical model of a dynamical system may appear in the form of phase portrait. The latter determines the "logic" of the behaviour of a separate motion. Let's construct the phase portrait of molecule

$$Z = X\sin[\xi(1 - X^{-2})]$$
$$Y = \frac{1}{2}Z\xi^{-1} + X^{-1}\left[\xi\left(1-X^{-2}\right)\right] \tag{14}$$

describes the phase trajectory. $Z \equiv z\Omega/V_0$ and $Y \equiv v/V_0$ are normalized distance and velocity, respectively. $X = \exp(\varepsilon t/2)$ is the parameter, which satisfy condition $(X^2/\xi) < 1$. The appearance of the phase trajectory is determined on the whole by equations (14) and corresponding initial conditions. Elimination of the parameter X is

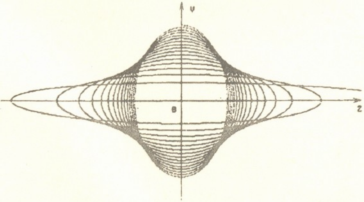


Fig. 2.

unpossible analytically, therefore using numerical simulation illustrate phase portrait for ε = 0.031 sec^{-1}, t=100 sec L = 100 sm, l = 12 sm, r_0 = 1 sm, R_0 = 6 sm, r_1 = 5 sm, $d_{eff} = 1.1\cdot 10^{-18}$ CGSE, $Q = 1.6\cdot 10^{-16}$ erg, $v_0 = 24\cdot 10^9$ sec^{-1} (see Fig.2.). The molecule passing from one ellipse to another aside the center (the equilibrium state), semimajor and semiminor axes are decreased forming the elliptical-logarithmical helix. So, parametrical interaction between the system and molecule leads to its deceleration. The phase portrait in three dimentional case is illustrated in Fig.3.

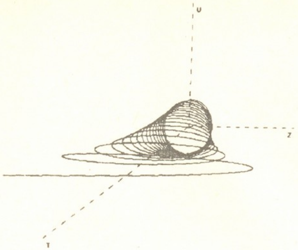


Fig. 3.

The condition of molecule deceleration one may find letting $\xi_0 \equiv \Omega/\varepsilon = 2.4$ which corresponds to the first root of the Bessel function of the first kind of the zeroth order, i.e. $J_0(\xi_0) = 0$. Equation (11) may be rewritten as:

$$Z(t) = \frac{\pi}{2}\frac{V_0}{\varepsilon}N_0(\xi_0)J_0(\xi_0 e^{-\varepsilon t}) = 0{,}8\frac{V_0}{\varepsilon}J_0(2.4e^{-\varepsilon t}). \tag{15}$$

At $\varepsilon t << 1$ expend exponent into series; keeping only first term, the molecule will stop in the piont $Z_0 \approx 0.8V_0/\varepsilon$ [5].

At great moments (including asymptotic $t \to \infty$) when condition

$$e^{\varepsilon t}/\xi >> 1\ (\xi > 1)$$

is fulfilled, equation (12) may then be written approximately

$$Z(t) = a + bt \tag{16}$$

where $a = V_0\left(\frac{\pi}{2\Omega\varepsilon}\right)^{1/2}\left[\sin\left(\xi - \frac{\pi}{4}\right) - \frac{2}{\pi}\cos\left(\xi - \frac{\pi}{4}\right)\left(\ln\frac{\xi}{2} + \gamma\right)\right]$,

$$b = V_0\left(\frac{2}{\pi\xi}\right)^{1/2}\cos\left(\xi - \frac{\pi}{4}\right),$$

γ is Euler's constant.

So, the molecule moves uniformly in straight lines inside the slowing-down system along the z axis. It has less velocity then at $t = 0$ moment $b < V_0$. "Effective" velocity is proportional to $b \sim \varepsilon^{1/2}$.

The solution of the Hill-Meissner equation (matrix presentation)

The equation (10) may be used as well as for the description of molecule motion inside a periodical, linear parametric system with lumped parameters in which the frequencies of time-varing parameters are proportionate. As it is well known, the behaviour of this system is described by linear differential equations with periodic coefficients.

Equation (14) one may split on two equations in the range $[0, T]$

$$\ddot{Z} + \Omega_1^2 e^{-\varepsilon t} Z = 0, \quad t \in \left[0, \frac{T}{2}\right] \tag{17}$$

$$\ddot{Z} + \Omega_2^2 e^{-\varepsilon t} Z = 0, \quad t \in \left[\frac{T}{2}, T\right] \tag{18}$$

Basic set of equations in the fundamental interval $0 \le t \le T$ is two linearly independent solutions: Bessel functions of the first and second kind of zeroth order. Following to [6,7] the solution of Hill-Meissner equation in the interval $[0, T]$ may be written in the matrix form

$$\begin{bmatrix} Z(T) \\ V(T) \end{bmatrix} = \frac{1}{W_0} \begin{bmatrix} J_0\left(\xi_2 e^{-\theta_1}\right) & \frac{1}{\varepsilon} N_0\left(\xi_2 e^{-\theta_1}\right) \\ -\Omega_2 J_0'\left(\xi_2 e^{-\theta_1}\right) e^{-\theta_1}, & -\frac{\Omega_2}{\varepsilon} N_0'\left(\xi_2 e^{-\theta_1}\right) e^{-\theta_1} \end{bmatrix} \times$$

$$\times \begin{bmatrix} J_0\left(\xi_1 e^{-\theta_1}\right), & \frac{1}{\varepsilon} N_0\left(\xi_1 e^{-\theta_1}\right) \\ -\Omega_1 J_0'\left(\xi_1 e^{-\theta_1}\right) e^{-\theta_1}, & -\frac{\Omega_1}{\varepsilon} N_0'\left(\xi_1 e^{-\theta_1}\right) e^{-\theta_1} \end{bmatrix} \times \tag{19}$$

$$\times \begin{bmatrix} -\frac{\Omega_1}{\varepsilon} N_0'(\xi_1), & -\frac{1}{\varepsilon} N_0(\xi_1) \\ \Omega_1 J_0'(\xi_1), & J_0(\xi_1) \end{bmatrix} \begin{bmatrix} 0 \\ V_0 \end{bmatrix} \equiv [M] \begin{bmatrix} 0 \\ V_0 \end{bmatrix},$$

where $\xi_1 = \dfrac{\Omega_1}{\varepsilon}$, $\theta_1 = \varepsilon \dfrac{T}{2}$.

Transitional second-order square matrix

$$[M] = \begin{bmatrix} A & B \\ C & D \end{bmatrix} \tag{20}$$

relates coordinate and velocity at the beginning and at the end of the fundamental interval.

Following Pipes [8], the solution of linear second-order differential equation with periodic coefficient may be introduced in the matrix form in the fundamental interval $0 \le t \le T$. Using Sylvester's theorem [9], the solution of the Hill-Meissner equation may be as stable as well as unstable. In particular, if

$$\left|\frac{A+D}{2}\right| \ge 1 \tag{21}$$

the solution is unstable; if

$$\left|\frac{A+D}{2}\right| < 1 \tag{22}$$

sollution has a stable character, such that Z and V are oscillate with increasing t.

Table 1

ε	T	θ_1	ξ_1	η_1	η_2	A	D	$\frac{A+D}{2}$
0.1	10	0.5	0.3	0.2	567	–0.02	2.9	1.5
0.5	10	2.5	0.07	0.005	2.1	0.06	–4.8	2.4
1	10	5	0.03	0.0002	0.6	–0.6	–0.5	0.5
1.5	10	7.5	0.02	0.00001	0.03	–0.6	0.002	0.3

Using conditions (21) and (22) from the Table 1 it follows that in the case of "weak"external electric field the molecule motion is unstable, contrary to the "strong" external field when molecule does not abandon the system.

Georgian Technical University

REFERENCES

1. *V.S.Letokhov*. Publ. CNRS, 217,1974, 127.
2. *G.V.Jandieri, Z.J.Surmanidze*. Proceedings of BSMA,1995.
3. *M.A.Eljashevich*. Atomic and Molecular Spectroscopy, Moscow, 1962 (Russian).
4. *P.K.Tien*. J. Appl. Phys., 25, 1954 1281.
5. *M.A.Miller*. Proc. Higher Educ. Inst. Radiophysics 5, 1962, 929.
6. *L.A.Pipes*. J. Appl. Phys., 24, 1953, 902.
7. *H.D'Angelo*. Linear Time-Varing Systems: Analysis and Synthesis, 1970.
8. Handbook of Mathematical Functions. Ed. by M. Abramowitz, I.A. Stegun, 1964.
9. *D.R.Merkin*. Introduction to the Theory of Stability Motion, Moscow, 1976. (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

PHYSICS

C.Tsiklauri, E.Khutsishvili, M.Mgaloblishvili

Microstructure and Microhardness of Gallium Arsenide

Presented by Corr. Member of the Academy G.Tsagareishvili, June 22, 1996

ABSTRACT. **Microstructure and microhardness of undoped gallium arsenide single crystals have been studied at room temperature. Dislocation structure of Ga(III) and As(III) faces was investigated. Polarity of microhardness is observed. Experimental data concerning the influence of annealing on the microhardness are reported.**

Gallium arsenide crystals are very important in semiconductor devices. This is connected with increasing applications of electronics in different fields of science and economy. Therefore crystal perfection of gallium arsenide gains special significance. Strict requirements for the crystal perfection of gallium arsenide require intense investigations in this direction. In the references [1-10] investigations of structure and microhardness of gallium arsenide have been described. Obtained data are not perfect and additional investigations are needed. Especially it concerns microhardness. Moreover the influence of thermal annealing on the microhardness is not investigated.

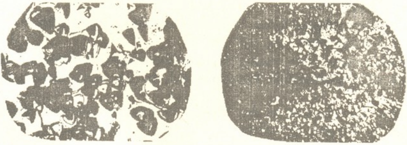


Fig.1 Dislocation structure of gallium arsenide on the (III) surface x 125

a) at surface edge b) at surface centre

For semiconductor devices and microschemes processing gallium arsenide substrates with [III] orientation are mainly used. In this paper we present the results of the experimental data of perfectionof the structure, microhardness and influence of thermalannealing (III) surfaces of galium arsenide single crystals on the microhardness. Dislocation structure has been studied by the method of chemical etching. Since dislocation density is lower than 10^6 dislocations/cm^2 we have observed dislocations on the microscope MIM-7 and Neophant. Microhardness was measured with PMT-3. Undoped n-type gallium arsenide crystals (with electron concentration $n \leq 6 \cdot 10^{16}$cm^{-3} and electron mobility 4200cm^2/v·sec) have been investigated.

After etching obtained surfaces can be classified into four types:

(1) relatively smooth surface with conical and pyramidal dislocation etch pits;

(2) relatively smooth surface without dislocation etch pits;

(3) surface completely covered with truncated pyramidal etch pits;

(4) surface covered with badly developed pyramidal etch pits.

(1) and (3) types of surfaces appear after etching on one side (Ga faces).

(2) and (4) types of surfaces appear after etching on the opposite side (As faces). Since thermal stress distribution is nonuniformed in the bulk of samples, consequently the same is observed for dislocation distribution. The thermal stress was expressed maximum at the centre and near the edge of sample. In accordance with stress change the maximum of dislocation density is observed in the centre and at the edge of sample surface (Fig. 1).

The estimated results of microhardness are listed in the Table. As seen from the Table polarity of microhardness is observed: Ga(III) face is harder than As(III) face, that is connected with dislocation velocites [9,10]. Thermal annealing of experimental samples was performed at 823k in hydrogen atmosphere for two hrs. Thermal annealing greatly influenced on the microhardness. After annealing effect of polarity is more pronounced. An increase of the microhardness is observed after annealing. This change is possibly connected with thermal stress being off after annealing.

Table

Microhardness of gallium arsenide H kg/mm^2

Ga face	As face note	Note
643±12	627±10	before annealing
735±13	674±14	after annealing at 823K 2 hr

F.Tavadze Institute of Metallurgy
Georgian Academy of Sciences

REFERENCES

1. *M.G.Milvidski.* Poluprovodnikovye Materialy v Sovremennoj Electronike. M., 1986.
2. *U.L.Pshenichnov.* Vijavlenie Tonkoj Strukturi Kristallov. M., 1974.
3. *M.G.Milvidski, B.B.Osvenski.* Strukturnie Defekti v Monokristallah Poluprovodnikov, M., 1984.
4. *R.L.Petrusevich.* F.T.T., **4**, 1962, 1378.
5. *R.L.Petrusevich.* Kristallografija, **8**, 1963, 243.
6. *J.W.Faust, J.A.Sagar.* J. Appl. Phys., **31**, 1966, 331.
7. *H.A.Scheli.* Z. Mett. Bard 48, Heft 4, 1957, 158.
8. *L.Berstein.* J. Elect. Soc., 109, 1962, 27.
9. *P.B.Hirsch, P.Pirour, S.G.Roberts, P.D.Warren.* Phil. Med. **50**, 1985,759.
10. *N.O.Uorren, S.R.Roberts, P.B.Hirsch.* Izv. AN SSSR, ser.fiz., **51**, 4, 1987, 812.
11. *K.Hilsum, A.Rous-inc.* Poluprovodniki Tipa $A^{III} B^{V}$, M., 1963.
12. *P.J.Stirland, W.Straugham.* Thin Solid Films, **31**, 1976, 139.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

GEOPHYSICS

V.Abashidze

Instrumental Studies of Differentiated Movements of the Blocks Underlying the Inguri HPS Dam

Presented by Academician B.Balavadze, September 9, 1996

Abstract. A complex analysis of geophysical instrumental observations has ascertained that there are no differentiated movements of the tectonic fault splitting the right-bank part of the Inguri dam basement into two structural blocks.

The right-bank fault of an upthrust-uplift type divides the rock underlying the Inguri HPS dam into two structural blocks [1]. Therefore the stability of this unique hydroengineering structure depends in many respects on the tectonic behaviour of this fault during the dam construction, filling and control of the reservoir.

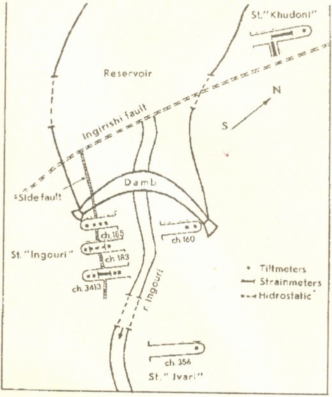


Fig.1 Lay-out of the tiltmetric and strainmetric observation stations in the Inguri HPS area.

Geological, structural and geomorphological studies revealed two stages - the Upper Pliocene-mid-Quaternary stage and the Upper Quaternary-Holocene stage in the geological and tectonic history of the area around the Inguri HPS dam. If in the former stage intensive differentiated movements of the structural blocks took place in the conditions of a general specific uplift, in the latter stage upward movements were still active, but differentiated movements weakened due to a general consolidation of the blocks [2], [3].

However geomorphological data cannot alone provide a unique quantitative solution of the problem about a mobility degree of the above-mentioned blocks. Hence high importance of the Inguri HPS dam makes it imperative to study this problem by instrumental geophysical methods [4].

Fig.1 shows the lay-out of the tiltmetric and strainmetric observation stations in the Inguri HPS area. As seen from the figure, thorough instrumental observations were organized in three adits built in close proximity of the fault. Highly sensitive tiltmeters installed here have been functioning since 1969, while from 1974 in adits Nos. 3413 and 183 additional observations have been organized by means of quartz strainmeters and hydrostatic level meters. These instruments are oriented to take measurements across the fault strike. Their measurement ranges are 22 and 35 metres respectively and they fix horizontal and vertical relative movements of the fault blocks. Even earlier, in 1971, to have additional information on the tectonic behaviour of the blocks of the right-bank fault, along with tiltmeters, strinometers with clock-type indicators designed as twisted-spring micrometers with a scale graduation of 1 μm were used. They were placed in two 1×1×1 m^3 recesses in adit No.185. These recesses were made in the filling material of the fault. The ends of the quartz tubes were grouted into the rock at both fault edges, while movements of the fault edges were registered in the filling material by means of measuring sensors installed at the free (unsecured) ends of the quartz tubes. The fault thickness at the observation site was 9 m; both fault edges had the well expressed glide planes.

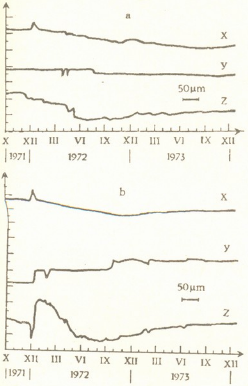


Fig.2 Deformometric data graphs
a) for the first glide plane of the fault;
b) for the second glide plane of the fault.

Movements of the fault edges were registered in three directions x, y and z. Readings of the twisted-spring micrometers were taken visually every two days. We began our deformometric observations in 1971 and continued them throughout more than two years. The results of the observation data analysis for all three directions are presented in the form of graphs in Fig.2 a and b for the first and second glide planes, respectively.

As seen from the graphs, changes in the readings, i.e. in displacements of the fault edges are observed only in the vertical (z) and horizontal (x) (across the fault strike) directions. No changes are noted in the direction (y) (along the fault) if we neglect individual sharp skew movements in all three directions due to explosions when excavating the dam foundation pit.

Thus, on the basis of deformometric observations, we come to a conclusion that there are no block movements along the fault in the inherited direction; the observed some microstrains in the vertical and horizontal directions indicate the mutual displacement of the fault edges relative to the filling material.

This conclusion was afterwards completely confirmed by long-term tiltmetric, strainmetric and hydrostatic observations. Namely, the results of long-term tiltmetric observations confirmed the conclusion of geostructural investigations about the absence of recent differentiated tectonic movements along the fault [5]. If differentiated movements took place, i.e., if the blocks moved relative each other in the fault plane in the inherited diretion, then the tilt vectors of the block would be directed along the fault, while the tilts themselves would be counterdirected. Actually, we did not observe such combinations of block movement directions in any of the dam construction stages. On the other hand, if such movement really took place, they would manifest themselves, in the first place, in the stages where the natural tectonic factors play the predominant role. But, just for these stages the tilt vectors of both structural blocks have the same stable south-eastward direction [6].

The results of deformometric and tiltmetric studies are also confirmed by data obtained due to observations of the fault zone by means of quartz strainmeters and hydrostatic level meters. Thus, an oscillatory character of displacements and linear deformations of the fault edges was found before the filling of the reservoir. The displacement velocities of the fault edges remained next to invariable in time, their mean value being 60-70 μm/yr.

Deformation processes were noted to become more active after the filling of the reservoir and especially at high water levels in the reservoir. A sharp water level control contributes to the development of relative vertical and horizontal displacements of the blocks underlying the dam. In that case nonelastic deformation develop in the form of residual deformation. When technogenous processes are quiet, displacements and deformations are minimal both in the fault and the entire dam basement.

Based on many-year complex observations, the daily norm of water level control in the reservoir was determined as follows: the rate of water level increase must not exceed 1 m/day, while the rate of lowering should be even less than that. This norm guarantees a safe operation of the Inguri HPS.

Georgian Academy of Sciences
Institute of Geophysics

REFERENCES

1. *G.M.Jigauri, A.K.Mastitski, S.B.Kereselidze.* Engineering and Geological Conditions of the Head Structure of the Inguri HPS and Jvari Water Reservoir (Engineering Report on the Inguri HPS, IV, book I), 1970, 262 (Russian).
2. *S.B.Kereselidze, V.Sh.Chumburidze..* Trudy Laboratorii Gidrologii i Inzhenerno-Geologich. Problem Gruzinsk. Tekhnich. Universitets 4, Tbilisi, 1972, 41-44 (Russian).
3. *A.K.Mastitski, G.M.Jigauri, S.B.Kereselidze.* J. Gidrotekhnicheskoye Stroitelstvo, M., 1974, 3-6.
4. *B.K.Balavadze, V.G.Abashidze, G.V.Mumladze.* Bull. Acad. Sci. Georgia, **52**, 1, Tbilisi, 1968, 51-55.
5. *A.K.Mastitski, S.B.Kereselidze, V.G.Abashidze.* Sbornik: Geologo-geophizich. Issledovan. v Raione Ingurskoi GES. Tbilisi, 1981, 147-164.
6. *B.K.Balavadze, V.G.Abashidze.* Tilts and Deformations of the Earth's Crust in the Inguri HPS Area. Tbilisi, 1985, 116 (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

GEOPHYSICS

G.Gunia, Z.Svanidze, L.Svanidze

The Determination of Metals' Microadmixtures in Atmospheric Air

Presented by Academician G.Svanidze, April 28, 1996

ABSTRACT. **The study of heavy metals in atmosphere using khelatforming fibrillar sorbents obtained by atomic-absorption method is discussed.**

The information of chemical components of atmospheric contaminant aerozol and change in time is of great practical interest. It deals with heavy metal group being the component of atmospheric dust. These researches have important meteorological, hygienic and economical aspects.

It is known [1] that heavy metals entering into the composition of atmospheric dust compose 10^{-2} - 10^{-12}% and frequently analytical test is too short. It is composed of the hundredth part of milligram up to several ones. All these make determination of aerosol composition rather difficult . In this case the classical methods of analytical chemistry are not effective and they can not answer strict demands of expressive mass analysis. The newest complex method of the research foresees the use of modernized sorbents and spectral apparatus guaranteeing exactness of the analysis.

Emissional spectral and atomic-absorptional methods [2] of analysis have particular significance giving possibility to carry out direct determination of the broad cycle of elements among the various substances.

For today there are several works devoted to the problem of study of heavy metal contents in the atmosphere [1,2]. Method of their emissive spectral analysis has been developed [3]. It was implemented in the former USSR government net work [4] against atmosphere pollution. For receiving more exact information on the components and concentrations of admixtures, for revealing the causes of their change, for evaluation of the meteorological conditions it is necessary to have the complete method of analysis in some cases. In connection with the mentioned method for investigation of the metal admixture in atmosphere, sensitive method should be worked out.

First of all the improvement of the methods of analysis should be directed to the increase of the exactness of the carried out determination and sensitivity, also for control of atmosphere pollution and for its simple use in practice.

As the researches showed, the basic dust components are elements such as: flint, aluminium, calcium, magnium and iron, the contents of which are changed from sample to sample. The alternating composition of the elements and their existence in great amount hinders the determination of admixture elements in little amount, its contents and components range being quite broad. Thus holding of the analysis with more high sensitivity requires preliminary activities for avoiding the background influence. The method of chemical treatment of samples can be used for this purpose, considering concentration beforehand and the removal of trace having elements from the basic components. The use of these methods provides high exactness of the analysis.

For today the number of methods using khelatforming sorbent for concentration and division of admixture elements is developed with more wide usage in the research practice for level of natural environmental pollution [5]. This essentially deals with the investigation of pollution of natural waters [6].

For quick and effective separation of metal microadmixtures from atmospheric air test we used khelatforming fibrillar sorbent POLIORGS VII M. We used it for the future determination of atomic-absorptional method.

Electroaspirator EA 822 is used for the test with air in Ig sorbent, which is located in 3.5 cm diameter glass funnel attached to it. This allowed to take the tests not only of the microadmixtures absorbed on atmospheric dust, but also of the smallest particles of metal microadmixtures dispersed in the air.

The sorbent is placed in 10ml 2M HNO_3 for 20 min for desorption, after which the sorbent is taken out from the solution. The composition of metal microadmixtures is determined by atomic-absorptional method. The optimal conditions are given in Table 1.

Table 1

The optimal conditions for absorptional determination of microelements

Parameters	Microelements							
	Cd	Zn	Cu	Pb	Co	Ni	Fe	Mn
Length of wave (nm)	228.8	213.9	324.8	288.3	240.7	232.0	248.3	279.5
Hole width	2	1	0.7	0.5	0.2	0.1	0.2	0.2
Speed of stripe	240	240	240	240	240	240	240	240
Flame	Acetylene - air							

As the amount of microadmixtures in the air test is very small, high requirements are put to the used sorbent. For this reason idle test is held; Ig sorbent is placed into 10 ml 2 $MHNO_3$ for 20 min.

Researches revealed, that maximum inaccuracy of our method does not exceed 10-15%.

The results of the analysis of atmospheric air tests taken in different districts of Tbilisi are given in Table 2.

Table 2

Composition of some metal microadmixtures in Tbilisi atmospheric air (mg l)

Sample	Microelements							
No	Cd	Zn	Cu	Pb	Co	Ni	Fe	Mn
1	0.31	0.21	0.06	0.07	0.81	0.56	0.02	0.01
2	0.56	0.18	0.05	0.09	0.75	0.65	0.01	0.05
3	0.93	0.16	0.07	0.03	0.82	0.72	0.03	0.02
4	0.26	0.15	0.04	0.06	0.78	0.48	0.09	0.05
5	0.29	0.13	0.08	0.08	0.87	0.48	0.05	0.07
6	0.64	0.2	0.08	0.06	0.76	0.88	0.09	0.08
7	0.50	0.15	0.04	0.03	0.74	0.48	0.03	0.08

As we see the present method is characterized with high sensitivity and gives us possibility to determine the wide spectrum of admixtures of metal elements, the contents of which in the air could be very insignificant.

The acquired results prove, that the discussed method is quite reliable and perspective for study and control of air basin pollution.

Georgian Technical University

REFERENCES

1. *G.S.Gunia.* Modern Problems of Atmospher Pollution. Obninsk, VNIIGMI - MTSD, 1978 (Russian).
2. *G.S.Gunia.* The Questions of Monitoring the Pollution of Atmospheric Air on the territory of Georgian SSR. L., 1985 (Russian).
3. *N.N.Aleksandrov, G.S.Gunia, A.I.Gunchenko, IU.I.Turkin.* Works GGO, 1974 issue 314.
4. The Guidance of Atmosphere Pollution Control. Edited by *M.E.Berliand, G.I.Sidorenko.* L., 1979 (Russian).
5. *Z.S.Svanidze, G.S.Gunia, G.V.Miasoedova, B.F.Miasoedov.* Works NII, L., 1986, issue 85(92) (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3 1996

ORGANIC CHEMISTRY

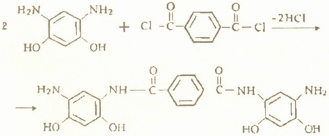
D.Tugushi, M.Gverdtsiteli

Algebraic Investigation of the Process of Condensation of o, o'-Disubstituted Aromatic Diamines with Chloranhydrides of Aromatic Dicarboxylic Acids

Presented by Corr. Member of the Academy L.M.Khananashvili, June 27, 1996

ABSTRACT. **Algebraic investigation of the process of condensation of o, o'-disubstituted aromatic diamines with chloranhydrides of aromatic dicarboxylic acids was carried out in terms of ANB - matrices method.**

Polybenzazoles are important representatives of polyheteroarylenes [1]. The first step of the model reaction of their synthesis is brought below:



(1)

Contiguity matrices of molecular graphs and their various modifications are widely used for algebraic characterization of molecules and their transformations in theoretical 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:

$$\begin{matrix} 1 & 2 & 3 \\ X & Y & V \end{matrix} \qquad \left\| \begin{matrix} Z_X & \Delta_{XY} & \Delta_{XV} \\ \Delta_{XY} & Z_Y & \Delta_{YV} \\ \Delta_{XV} & \Delta_{YV} & Z_V \end{matrix} \right\| \qquad (2)$$

where Z_X, Z_Y, Z_V are atomic numbers of X, Y, 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 must be mentioned, that for large molecules calculations are very labour-consuming. We have eleborated a simple model, which reproduces the specific character of system and at the same time is less labour-consuming; o,o' - dihydroxisubstituted aromatic diamine can be written as:

$$A - NH_2, \qquad (3)$$

where A represents the molecule of o,o'-dihydroxisubstituted aromatic diamine without NH_2 group. Corresponding modernized ANB - matrix ($\widetilde{ANB}$) has a form:

$$\begin{Vmatrix} Z_A & 1 & 0 & 0 \\ 1 & 7 & 1 & 1 \\ 0 & 1 & 1 & 0 \\ 0 & 1 & 0 & 1 \end{Vmatrix}, \tag{4}$$

where

$$Z_A = \sum Z_a \tag{5}$$

and Z_a are the atomic numbers of chemical elements containing A fragment.
Chloranhydride of aromatic dicarboxylic acid can be written as:

$$Cl - B - Cl, \tag{6}$$

where B - represents the molecule of chloranhydride of aromatic dicarboxylic acid without two chlorine atoms. Corresponding modernized ANB - matrix ($A\tilde{N}B$) has a form:

$$\begin{Vmatrix} 17 & 1 & 0 \\ 1 & Z_B & 1 \\ 0 & 1 & 17 \end{Vmatrix}, \tag{7}$$

where

$$Z_B = \sum Z_b \tag{8}$$

and Z_b are the atomic numbers of chemical elements containing B fragment.
The model notation of (1) process has a form:

$$A - NH_2 + Cl - B - Cl + A - NH_2 \rightarrow$$
$$\rightarrow A - NH - B - NH - A + 2HCl \tag{9}$$

The matrix notation of (9) process has a form:

$$\begin{Vmatrix} 65 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 1 & 7 & 1 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 1 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 1 & 0 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 17 & 1 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 1 & 68 & 1 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 1 & 17 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 65 & 1 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 7 & 1 & 1 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 1 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 0 & 1 \end{Vmatrix} \longrightarrow$$

$$\begin{Vmatrix} 65 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 1 & 7 & 1 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 1 & 1 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 1 & 0 & 68 & 1 & 0 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 1 & 7 & 1 & 1 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 1 & 1 & 0 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 1 & 0 & 65 & 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 1 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 17 & 0 & 0 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 1 & 1 \\ 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 0 & 1 & 17 \end{Vmatrix} \qquad (10)$$

Calculations show, that the value of determinant of $A\tilde{N}B$ - matrices describing (1) process increased. It is algebraic criterion (in terms of $A\tilde{N}B$ - matrices method) for this process.

I.Javakhishvili Tbilisi State University

REFERENCES

1. *A.L.Rusanov, D.S.Tugushi, V.V.Korshak.* Progress in Chemistry of Polyheteroarylenes. Tb., 1988.
2. *G.Gamziani.* Mathematic Chemistry. Tb., 1990.
3. *G.Gamziani, M.Gverdtsiteli.* Phenomenon of Isomery from Point of View of Mathematic Chemistry. Tb., 1992.
4. *M.Gverdtsiteli.* Principles of Nomenclature of Chemical Compounds., Tb., 1983.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

ORGANIC CHEMISTRY

T.Guntsadze, M.Buachidze, M.Gverdtsiteli

Algebraic Investigation of the Process of Interaction of Acetylenic Carbinols with $HGe(C_2H_5)_3$

Presented by Corr. Member of the Academy D.Ugrekhelidze, June 27, 1996

ABSTRACT. **Algebraic investigation of the process of interaction of acetylenic carbinols with $HGe(C_2H_5)_3$ was carried out in terms of ANB - matrices method. Its diagonal elements represent atomic numbers of chemical elements, whereas nondiagonal ones - the multiplicities of chemical bonds.**

The process of interaction of primary and secondary vinylacetylenic carbinols with $HGe(C_2H_5)_3$ can be written as [1]:

$$\underset{\displaystyle |\atop OH}{CH_3-CH}-C\equiv C-CH=CH_2+HGe(C_2H_5)_3\rightarrow CH_3-\underset{\displaystyle OH}{\overset{}{CH}}-\overset{Ge(C_2H_5)_3}{C}=CH-CH=CH_2 \qquad (1)$$

Contiguity matrices of molecular graphs and their various modifications are widely used for algebraic characterization of molecules and their transformations in theoretical organic chemistry [2, 3]. One type of such matrices 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:

$$\begin{matrix} 1 & 2 & 3 \\ X & Y & V \end{matrix} \qquad \left\|\begin{matrix} Z_X & \Delta_{XY} & \Delta_{XV} \\ \Delta_{XY} & Z_Y & \Delta_{YV} \\ \Delta_{XV} & \Delta_{YV} & Z_V \end{matrix}\right\| \qquad (2)$$

where: Z_X, Z_Y, Z_V are atomic numbers of X, Y, 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 must be mentioned, that for large molecules calculations are very labour - consuming. We have elaborated a simple model, which reproduces the specific character of system and at the same time is less labour - consuming. For carbinol the model has a form:

$$\begin{matrix} 1\;\;2\;\;3\;\;4 \\ A-C\equiv C-B \end{matrix} \qquad (3)$$

where: A - represents $CH_3CH(OH)$ fragment; B – $CH = CH_2$ fragment Δ.
Corresponding modernized ANB - matrix ($A\tilde{N}B$) has a form:

$$\left\|\begin{matrix} Z_A & 1 & 0 & 0 \\ 1 & 6 & 3 & 0 \\ 0 & 3 & 6 & 1 \\ 0 & 0 & 1 & Z_B \end{matrix}\right\| \qquad (4)$$

where:

$$Z_A = \sum Z_a \quad (5)$$

$$Z_B = \sum Z_b \quad (6)$$

(5) and (6) represent the sums of atomic numbers of chemical elements which A and B fragments contain.

For $HGe(C_2H_5)_3$ the model has a form:

$$D - H \quad (7)$$

where D - represents $Ge(C_2H_5)_3$ fragment. Corresponding $\widetilde{ANB}$ matrix has a form:

$$\begin{Vmatrix} Z_D & 1 \\ 1 & 1 \end{Vmatrix}, \quad (8)$$

where:

$$Z_D = \sum Z_d \quad (9)$$

represents the sum of atomic numbers of chemical elements which D fragment contains.

So, the model notation of (1) process has a form:

$$A - C \equiv C - B + D - H \rightarrow A - CD = CH - B \quad (10)$$

The matrix notation of (10) process has a form:

$$\begin{matrix} 25 & 1 & 0 & 0 & 0 & 0 \\ 1 & 6 & 3 & 0 & 0 & 0 \\ 0 & 3 & 6 & 1 & 0 & 0 \\ 0 & 0 & 1 & 15 & 0 & 0 \\ 0 & 0 & 0 & 0 & 83 & 1 \\ 0 & 0 & 0 & 0 & 1 & 1 \end{matrix} \longrightarrow \begin{matrix} 25 & 1 & 0 & 0 & 0 & 0 \\ 1 & 6 & 2 & 0 & 1 & 0 \\ 0 & 2 & 6 & 1 & 0 & 1 \\ 0 & 0 & 1 & 15 & 0 & 0 \\ 0 & 1 & 0 & 0 & 83 & 0 \\ 0 & 0 & 1 & 0 & 0 & 1 \end{matrix} \quad (11)$$

Calculations show, that during (11) transformation the value of determinant of $\widetilde{ANB}$ matrices increased and it is algebraic criterion for this process (in terms of ANB - matrices method).

Analogous algebraic criterion was found for diacetylenic glycols [5].

I.Javakhishvili Tbilisi State University

REFERENCES

1. *I.M.Gverdtsiteli, T.P.Guntsadze, A.D.Petrov.* PAS USSR, **157**, 3, 1964.
2. *G.Gamziani.* Mathematic Chemistry. Tb., 1990.
3. *G.Gamziani, M.Gverdtsiteli.* Phenomenon of Isomery from Point of View of Mathematic Chemistry. Tb., 1992.
4. *M.Gverdtsiteli.* Principles of Nomenclature of Organic Compounds., Tb., 1983.
5. *I.M.Gverdtsiteli, M.A.Buachidze.* PAS USSR, **158**, 1, 1964.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

ORGANIC CHEMISTRY

M.Gverdtsiteli, G.Chikvinidze, I.Gverdtsiteli

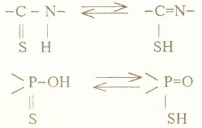
Algebraic Characterization of Thion-Thiol Triad Prototropic Tautomerism

Presented by Corr. Member of the Academy L.Khananashvili, June 3, 1996.

ABSTRACT. **The algebraic method of organic compounds notation in the form of square symmetric matrices is considered. Their diagonal elements' represent the ordinal numbers of chemical elements, whereas nondiagonal ones - the multiplicity of chemical bonds (ONB - matrices). Algebraic characterization of the thion - thiol triad prototropic tautomerism is given in terms of this method.**

Algebraic characterization of thion-thiol triad prototropic tautomerism is given in terms of ANB - matrices method. Their diagonal elements represent the ordinal numbers of chemical elements, whereas nondiagonal ones - the multiplicities of chemical bonds.

In chemistry of sulphur-organic compounds, nitrous and phosphor thion - thiol tautomerism is one of the most interesting processes [1]. This process has a scheme:



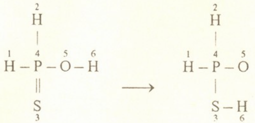
Contiguity matrices and their various modifications are efficiently used in modern theoretical organic chemistry for characterization of molecules and their transformations [2,3]. One type of such matrices are ANB - matrices - ther diagonal elements represent atomic numbers of chemical elements, whereas nondiagonal elements - the multiplicities of chemical bonds [4].

For arbitrary ABC molecule ANB - matrix has a form:

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

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, and C (if ABC molecule is not cyclic, Δ_{AC} is equal to 0).

The modelling reactions of thion - thiol tautomerism and their notation in matrix form are brought below:



1	0	1	0	0	0		1	0	1	0	0	0
0	16	2	0	0	0		0	16	1	0	1	0
1	2	6	1	0	0	→	1	1	6	2	0	0
0	0	1	7	1	1		0	0	2	7	0	1
0	0	0	1	1	0		0	1	0	0	1	0
0	0	0	1	0	1		0	0	0	1	0	1

1	0	0	1	0	0		1	0	0	1	0	0
0	1	0	1	0	0		0	1	0	1	0	0
0	0	16	2	0	0	→	0	0	16	1	0	1
1	1	2	15	1	0		1	1	1	15	2	0
0	0	0	1	8	1		0	0	0	2	8	0
0	0	0	0	1	1		0	0	1	0	0	1

Let consider the expression:

$$\Delta_r = \Delta_f - \Delta_i$$

where: Δ_i- is the value of determinant of ANB - matrix for initial tautomer; Δ_f - for final tautomer; Δ_r- change of value in the result of tautomeric transformation.

In the case of nitrous thio - thiol tautomerism $\Delta_r = 20$; for phosphor thion-thiol tautomerism $\Delta_r = 80$.

Thus, the algebraic criterion for both tautomeric processes (in terms of this method) is the increasing of values of determinants of ANB - matrices.

I. Javakhishvili Tbilisi State University

REFERENCES

1. *Z.Slanina*. Theoretical Aspects of Phenomenon of Isomery in Chemistry. M., 1984.
2. *G.Gamziani*. Mathematic Chemistry. Tb., 1990.
3. *G.Gamziani, M.Gverdtsiteli*. Phenomenon of Isomery from Point of View of Mathematic Chemistry. Tb., 1992.
4. *M.Gverdtsiteli*. Principles of Nomenclature of Organic Compounds. Tb., 1983.

PHYSICAL CHEMISTRY

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

Determination of Discrimination Indices of Capillary Column Applied for Studies on Sorbent-Sorbate Adsorption Interaction

Presented September 2, 1996

ABSTRACT. **Discrimination indices of quartz aluminized columns of different diameters for n-alkanes have been studied.**

n-dekan peak area is taken for a criterion. It is shown that quartz aluminized column with a diameter of 0.53 mm and 3mm thickness of the film gives good results in the study of physico-chemical interactions of sorbent-sorbate.

The results of determination of the discrimination indices of n-alkanes mixture are given in Tables 1a, b. The criterion for the quantity of the introduced sample is assumed to be the area under the peak of n-decane. For evidence the tests have been grouped in the order of reduction of the sample volume, entering the column with gas from the carrier.

It is evident from the tables that the distribution of the ratio of the quantities of sample components,entering the column from the evaporator takes place at random and a relative standard deviation Sr of adjacent members of the homological series in the mixture of the limited content is 10% for C_{11}....C_{14} and 3.4% for the adjacent members of the homological series.

Table 1b

C_{12}-Peak Area	№ of analysis	Relation of Peak Areas			
		C_{14}/C_{11}	C_{14}/C_{12}	C_{12}/C_{14}	
36109	6	1.70	1.26	1.35	
22888	8	1.50	1.15	1.30	$\overline{X} = \frac{\sum X_i}{n}$
18039	7	1.87	1.43	1.30	
15476	4	1.58	1.12	1.41	$\delta = \sqrt{\frac{(x_i - x)^2}{n-1}}$
10346	5	1.80	1.32	1.36	
	$\overline{X}$	1.69	1.26	1.34	
	δ	0.152	0.127	0.046	$S_r = 100 \frac{\sigma}{\overline{X}} \%$
	S_r	9.0	10.1	3.4	

An increase of discrimination has been observed in the mixture of n-alkanes with much more components in it

Table 1b

C_{12}-Peak Area	№ of analysis	Relation of Peak Areas			
		C_{14}/C_{11}	C_{14}/C_{12}	C_{12}/C_{14}	C_{14}-C_6
48107	3	1.36	1.25	1.09	57.9
38984	4	1.36	1.25	1.09	51.6
38207	4	1.17	0.99	1.18	36.8
29100	2	1.24	1.08	1.14	39.7
19892	5	1.56	1.31	1.19	44.9
	$\overline{X}$	1.34	1.18	1.14	46.2
	δ	0.15	0.13	0.05	8.64
	S_r	11.09	11.46	4.19	18.7

Thus for C_{11}/C_6 it equals to 18.7% at confidence level of 50% and preservation stochastic dependence, having no correlation with the volume of injected sample.

In an effort to reveal the influence of the chemical structure of substances on the discrimination of the components, the mixtures of n-alkanes C_5......C_{10} with aromatic compounds such as benzene, toluene, ethylbenzene, m- , p- and o-xylols are being' introduced. The determined relation of peak areas arranged in n-C_{10} peak area order are presented in Table 1c.

Table 1c

C_{10} Peak Area	No of analysis	Relation (ship) of peak areas					
		C_{10} / benzene	C_{10} / C_8	C_9 / toluene	C_{10} / toluene	C_{10} / ethylbenzene	O-xylene / toluene
168920	14	2.59	6.18	4.68	1.20	-	3.19
144036	9	7.13	2.72	1.67	2.77	1.59	2.11
103984	11	6.03	2.60	1.58	2.56	1.48	2.09
72064	10	7.03	2.78	1.65	2.77	1.59	2.12
64865	12	7.64	2.25	1.63	3.03	1.75	2.14
46947	8	7.10	2.81	1.64	2.76	1.61	2.11
36425	6	9.21	3.01	1.88	3.22	1.66	2.42
35693	7	8.64	2.96	1.81	3.05	1.63	2.36
32578	1	7.02	2.75	1.66	2.70	1.56	2.14
12946	2	5.54	2.81	1.70	2.72	1.61	2.02
11388	13	4.99	2.90	1.78	3.16	1.65	2.32
general	$\overline{X}$	6.63	3.13	1.97	2.72	-	2.28
selection	δ	1.82	1.02	0.90	0.55	-	0.32
n = 11	S_r	27.4	32.5	45.8	20.1	-	14.2

It should be noted that according to the rule of "three sigmas" general population comprises all 11 correct variants, due to it the results are no failure and can be considered as a characteristic phenomenon. In this case the highest' extent of discrimination is due to the difference in the structure of sorbate molecules (Sr of the relation C_9/toluene amounts to 45.8%) and in the second place to the difference in sizes of molecules of the compared components.

At the same time the absolute values of the variant of the first line differs to such extent from the rest of the selected data that we have assumed it possible to consider the selected population without taking into account the variant of analysis No14; from here it follows that the highest value of discrimination Sr=18.4% is observed for carbohydrates with both different structure and much more difference in molecular masses that is evidenced by the reducing order: C_{10}/benzene>C_{10}/toluene>o-xylol/toluene>C_9/toluene>C_{10}/ethylbenzene built up according to the data of Table 1d (continuation of Table 1c):

Table 1d

Selected population n = 10 (without-N14)							
	$\overline{X}$	7.03	2.83	1.70	2.87	1.61	2.19
	δ	1.296	0.126	0.094	0.223	0.070	0.125
	S_r	18.4	4.5	5.5	7.7	4.3	5.7

If validity of all variants will be assumed, the conclusion suggests that overload may lead to considerable show of the sample discrimination, especially in case of relatively high releases.

The carried out evalution shows that capillary columns of large diameter must have less discrimination of the components in comparison with the "classical" ones due to release reduction.

The experimental test of the column with the diameter of 0.53mm showed that for C_9/toluene at the change of peak area C_{10} in 15 times, a relative standard deviation Sr = 5.5% at the release 1:2:5, pointing to the discrimination absence under almost non-release condition indicates direct way to the application of broad capillary columns for physico-chemical investigation in sorbent-sorbate system.

Thus physico-chemical interactions in sorbent sorbate system can be determined by the highest correctness using aliminized quartz capillary columns of a large diameter (0.53mm) with a thick (up to 3 mkm) film of SLPH due to their high velocity, absence of activity and discrimination, and a unique strength.

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

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

ELECTROCHEMISTRY

R.Kvaratskhelia, H.Kvaratskhelia

Voltammetry of Cr (6+) in Acidic Dichromate Solutions at the Solid Electrodes

Presented by Corr. Member of the Academy L.Japaridze, July 7, 1996

ABSTRACT. **It has been shown that the predominant form of existence of Cr (6+) in the $H_2Cr_2O_7$ aqueous solutions in neutral (0.1M $NaClO_4$) and acidic (0.1N H_2SO_4) media is $HCrO_4^-$ ion. The main kinetic parameters of the Cr (6+) electroreduction were calculated. Nature of the Cr (6+) waves was explained.**

The present work has been carried out within the framework of the study of acids eletroreduction with the electrochemically active anions.

The study of electrochemical behaviour of Cr (6+) in the aqueous solutions of $H_2Cr_2O_7$ was carried out by the methods of voltammetry at the rotating disk electrodes and chronovoltammetry at the stationary electrodes from high-purity Sn, Cu and Cu-Hg in the sealed cell with a pure helium atmosphere (in the case of the above-mentioned metals the most clear picture of process is observed).

Technique of a preparation of the electrodes for the measurements has been described in [1]. The background electrolyte used in the study - $NaClO_4$ has been twice recrystallized from a bidistillate and then calcinated at 190-200°C for several days. Twice distilled H_2SO_4 has been also used in the study. Twice recrystallized from a bidistillate $K_2Cr_2O_7$ has been used as Cr (6+) containing compound. 0.1M $K_2Cr_2O_7$ and 0.1M H_2SO_4 were used for the preparation 0.1M solution of $H_2Cr_2O_7$. The saturated calomel electrode has been used as a reference electrode. All the measurements have been carried out at 20°C.

Taking into account the complex equilibriums existing between the various forms of Cr (6+) in the aqueous solutions with pH 0-6 we have examined the problem of nature of the reducing particle in this medium. The following equilibriums have been considered here:

$$2H_2CrO_4 \rightleftarrows Cr_2O_7^{2-} + H_2O + 2H^+ \qquad (1)$$

$$H_2CrO_4 \rightleftarrows HCrO_4^- + H^+ \qquad (2)$$

$$Cr_2O_7^{2-} + H_2O \rightleftarrows 2HCrO_4^- \qquad (3)$$

$$HCrO_4^- \rightleftarrows CrO_4^{2-} + H^+ \qquad (4)$$

$$Cr_2O_7^{2-} + H_2O \rightleftarrows 2CrO_4^{2-} + 2H^+ \qquad (5)$$

Basing on the quantitative data of these equilibriums (Pourbaix diagrams with a supplement [2]) we have obtained the equations connecting the ratios of the concentrations of the participants of the equilibriums (1)-(5) with pH and the total concentrations of the Cr (6+) containing compound (in regard to CrO_4^{2-}) in the solution:

$$x_1^2 - \frac{10^{-0.18-2pH}}{C_{tot}} x_1 - \frac{2\cdot 10^{-0.18-2pH}}{C_{tot}} = 0 \qquad (6)$$

$$x_2 = 10^{0.75 - pH} \qquad (7)$$

$$x_3^2 - \frac{0.0209}{C_{tot}} x_3 - \frac{0.0418}{C_{tot}} = 0 \qquad (8)$$

$$x_4 = 10^{6.45 - pH} \qquad (9)$$

$$x_5^2 - \frac{10^{-14.59+2pH}}{C_{tot}} x_5 - \frac{2\cdot 10^{-14.59+2pH}}{C_{tot}} = 0 \qquad (10)$$

In these equations:

$$x_1 = \frac{[H_2CrO_4]}{[Cr_2O_7^{2-}]};\ x_2 = \frac{[H_2CrO_4]}{[HCrO_4^-]};\ x_3 = \frac{[HCrO_4^-]}{[Cr_2O_7^{2-}]};\ x_4 = \frac{[HCrO_4^-]}{[CrO_4^{2-}]};\ x_5 = \frac{[CrO_4^{2-}]}{[Cr_2O_7^{2-}]}.$$

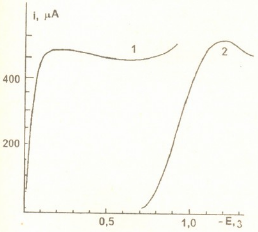


Fig.1 The dependence of the ratios of concentrations of the various Cr (6+) forms on pH and C_{tot} (in the cases of x_1 and x_5 C_{tot} = 0.002M)

The calculation results made by the equations (6)-(8) and (10) are presented in Fig.1. The obtained data show that the predominant form of Cr (6+) in the used $H_2Cr_2O_7$ solutions (C_{tot} = 0.001 - 0.005M) in neutral (0.1M $NaClO_4$) and acidic (0.1N H_2SO_4) media is $HCrO_4^-$ ion.

In all the used solutions and at all the used electrodes the clear voltammetric picture of Cr (6+) reduction is observed. In the neutral background solutions it forms two waves at the Sn and Cu - Hg electrodes and one wave at copper electrode, but in acidic medium - one wave at the Cu and Cu - Hg electrodes (in Fig.2 the case of Cu electrode is shown). It is necessary to note that the values of the limiting currrents of Cr (6+) first wave in 0.1M $NaClO_4$ increase appreciably with a rise of the H_2SO_4 concentration in solution. In the table the values of the kinetic parameters of the process-half-wave potential $E_{1/2}$ and rate constant k_0 are presented (the latter have been calculated as a result of analysis of the chronovoltammograms parameters).

The reduction process in all the cases is carried out in the diffusion regime. It is confirmed by the analysis results of the limiting currents

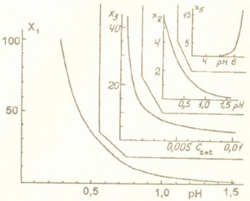


Fig.2 The voltammograms of Cr (6+) at the copper electrode. 10^{-3}M $H_2Cr_2O_7$; 980 r.p.m. 1 - 0.1N H_2SO_4; 2 - 0.1M $NaClO_4$

The main kinetic parameters of electroreduction of Cr(6+)

Electrode	0.1 M $NaClO_4$		0.1N H_2SO_4	
	$E_{1/2}$, V	k_0, cm/s	$E_{1/2}$, V	k_0, cm/s
Sn				
I wave	-0.66	$6.0\cdot10^{-8}$	–	–
II wave	-1.34	–	–	–
Cu-Hg				
I wave	-0.07	–	-0.02	$2.5\cdot10^{-4}$
II wave	-1.42	–	–	–
Cu	-0.93	–	-0.03	$1.1\cdot10^{-4}$

with the aid of Levich's equation.

Considering the fact that at the cathode the following reaction takes place:

$$HCrO_4^- + 3e + 7H^+ \rightarrow Cr^{3+} + 4H_2O \qquad (11)$$

we have analyzed the process of Cr (6+) reduction with the aid of equation suggested by us [3] describing the kinetics of acids electroreduction with the electrochemically active anions. In the given case this equation has the following form:

$$\frac{i_d^{HCrO_4^-}}{i_d^{H_3O^+}} = 3\frac{C_{HCrO_4^-}}{C_{H_3O^+}}\left(\frac{D_{HCrO_4^-}}{D_{H_3O^+}}\right)^{2/3} \qquad (12)$$

We calculated the value of $D_{HCrO_4^-}$ with the aid of Levich's equation considering the values i_d of Cr (6+) in 0.1M H_2SO_4; it is equal to $0.54\cdot10^{-5}$ cm^2/s. The ratio of the limiting diffusion currents of $HCrO_4^-$ and H_3O^+ calculated from the equation (12) is equal to 0.449. Thus, $HCrO_4^-$ ion in the neutral solutions containing the acidic forms of Cr (6+) must give one wave corresponding to a proton-donor action of H_3O^+ ions. Considering this conclusion, it is possible to affirm that the wave of Cr (6+) in 0.1N H_2SO_4 and I wave in 0.1M $NaClO_4$ correspond to one process - reduction of $HCrO_4^-$ ion in accordance with reaction (11). But the limiting currents of these waves are appreciably different. This fact (considering increase of the height of the first wave in 0.1M $NaClO_4$ with the rise of H_2SO_4 concentration in solution) may be explained as follows. Owing to the consumption of H_3O^+ ions during reaction (11) at a low content of these ions in the 10^{-3} M $H_2Cr_2O_7$ solution in 0.1 M $NaClO_4$ the equilibrium (4) in the reaction zone while registering voltammograms shifts in the direction of the electrochemically inactive (in the given region of potentials) CrO_4^{2-} ions. The content of $HCrO_4^-$ ions in the above mentioned zone decreases.

Addition of H_2SO_4 preventing this equilibrium shift causes increase of the limiting current of Cr (6+). As to the second wave of Cr (6+) in 0.1M $NaClO_4$ (and the wave in the case of copper electrode) located in the more negative region of potentials, it is necessary to consider increase of pH in the electrode – adjoining layer causing appreciable shift of the equilibriums (4) and (5) in the direction of CrO_4^{2-} ions. The above mentioned waves correspond to reduction of these ions:

$$Cr\,O_4^{2-} + 3e + 8H^+ \rightarrow Cr^{3+} + 4H_2O \qquad (13)$$

In the region of second wave potentials the formation of the film at the electrode surfaces is observed. It corresponds to the Cr (3+) oxide - hydroxide forms.

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

REFERENCES

1. *R.K.Kvaratskhelia, T.Sh.Machavariani.* Elektrokhimiya, **20**, 3, 1984, 303.
2. Spravochnik Khimika, t III, Moskva-Leningrad, 1964 (Russian).
3. *R.K. Kvaratskhelia, M.G. Zhamierashvili, H.R. Kvaratskhelia.* Elektrokhimiya, **28**, 12, 1992, 1869.
4. *R.K. Kvaratskhelia, M.G.Zhamierashvili, H.R. Kvaratskhelia.* Bull. Georg. Acad. Sci., **145**, 3, 1992, 569.

ᲡᲐᲥᲐᲠᲗᲕᲔᲚᲝᲡ ᲛᲔᲪᲜᲘᲔᲠᲔᲑᲐᲗᲐ ᲐᲙᲐᲓᲔᲛᲘᲘᲡ ᲛᲝᲐᲛᲑᲔ, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

GEOLOGY

G.Gongadze

Caucasian Echinoids on the Maastrichtian and Danian Boundary

Presented by Academician I.Gamkrelidze, July 18, 1995

ABSTRACT. **On the boundary of Maastrichtian and Danian in the Caucasus cutting down of Echinoids is clearly noticeable. It is the result of Mesozoic extinction. This is due to the fact that Danian-Paleocene Echinoids represent the impoverished part of the Upper Cretace complex.**

We think it is not right to draw the Cretace-Paleogene boundary between Maastrichtian and Danian by Echinoids. This boundary should be drawn between Danian and Paleocene or between Paleocene and Eocene.

Maastrichtian/Danian boundary occupies a special place in the evolutional development of organic world. Across the boundary "a great extinction" leading to the extinction of half of genera and three-fourth of species of animals and plants, had happened. The "grands" of Mesozoic era, such as ammonites, belemnits,inoceramids, dinosaurs, also rudists, pterosaurs, mosasaurs and some others proved to be unable to cross this boundary. Planctonic foraminifers, brachiopods, corals, echinoids and other organisms sustained great losses.

According to the analyses of many groups of organisms during the Cretaceous-Paleogene, extinction was going on gradually and not catastrophically. At the same time the boundary is crearly expressed between Maastrichtian and Danian, where the large taxons of fauna considerably reduced or completely died out during a short period of time.

What do Caucasian Late Cretaceous and Paleogene echinoids show in this aspect? According to our data all stages of the Caucasian Upper Cretaceous (also Danian) are more or less well characterized by echinoids. In this respects Campan-Maastrichtian-Danian are important (first of all Maastrichtian). It's very interesting that from 19 Maastrichtian genera only 6 passed Maastrichtian/Danian boundary: *Echinocorys, Galeaster, Cyclaster, Coraster, Orthaster, Homoeaster*. At the end of Maastrichtian: *Conulus, Galerites, Catopygus, Oolopygus, Cardiaster, Offaster, Hemipneustes, Stegaster, St.*(*Seunaster*), *Guettaria, Pseudoffaster, Micraster* (*Isomicraster*) were extinct.

As we noted in the Caucasian region Maastrichtian/Danian boundary is clearly expressed by echinoids. The coefficient of extinction (of genera) is 64 per cent. Conulidae, galeritidae, nucleolitidae were unable to cross the Maastrichtian/Danian boundary. Highly suffered and not completely extinct holasteroids, from which four genera reached the Maastrichtian/Danian boundary and the only genus (*Echinocorys*) passed to Danian. We have the same picture of stegasteroids. From five genera only one crosses this boundary. In the Caucasian Danian seas the Basseaster appeared for the first time. At the same time 5 new genera appeared, from which *Garumnaster* and *Jeronia* were extinct at the end of Danian; *Isopneustes* reached the Mons, *Pseudogibbaster* reached the end of Paleocene and *Neoglobator* - reached the Middle Eocene. It is remarkable that all these genera besides *Basseaster* and *Garumnaster* have a well expressed Upper Cretaceous appearance. The Upper Cretaceous appearance of the Caucasian Danian echinofaunas is increasing because there are numerous representatives of genera from Turonian and from the following stages: *Echinocorys* (13 species), *Orthaster* (5 species), *Coraster* (4 species).

In Caucasian Lower Paleocene there are known 9 genera: *Neoglobator*, *Echinocorys*, *Galeaster*, *Pomaster*, *Pseudogibbaster*, *Isopneustes*, *Coraster*, *Orthaster*, *Homoeaster*. There are also a few echinoids in Upper Paleocene: *Neoglobator*, *Echinocorys*, *Pomaster*, *Pseudogibbaster*, *Isaster*, *Brissopneustes*, *Coraster*, *Orthaster*. Enumerated genera show, that Paleocenian complex of Caucasian echinofauna is impoverished Danian complex, but both of them are the part of Upper Cretaceous united type of echinofauna.

There are 7 genera in Caucasian Eocene, from which except two *Neoglobator*, *Brissopneustes*, others *Echinolampas*, *Conoclypus*, *Pericosmus*, *Parabrissus*, *Meskhetia* express very well *Cenozoic* type of echinofauna.

As it was noted, the extinction of bios is graduall, not catastrophic, but in our opinion a big catastrophe happenning on the earth at the end of Maastrichtian (remember "Impact hypothesis") could change the ecological conditions of sea and continental organisms. It made fast or finished the extinction of certain groups. Other groups seriously suffered after general catastrophe, they resisted the whole extinction.

In relation to the Mesozoic bios extinction problem, there is the second, also important and at the same time very vexed question about Cretaceous/Paleogene (Mesozoic/Cenozoic) boundary. From the "great extinction" event, more natural seems to draw this boundary between Maastrichtian and Danian, though we think this problem is not so simple. We have no possibility to analyse in detail this complex problem. We can only notice, that to establish the boundaries between stratigraphic units both general geological factors and paleontological (faunistic-floristic) data have to be taken into consideration. On this concrete occasion we should consider the aforementioned question basing on the statistics about echinofauna.

Echinoids are widespread in Maastrichtian-Danian deposits. Late Cretaceous epoch is fluorishing time for this group, when Upper Cretaceous type of echinofauna was formed basically. Maximum of development comes to the Maastrichtian century. At this time a great many families and genera with great variety of species existed. However this variety of taxons were reduced sharply at the beginning of Danian because of aforenoted extinction of biota. Exactly because of this extinction we can explain the obvious regress in development of echinofauna in the Danian-Paleocenian centuries. In Eocene, on the contrary, the new evolutionary "burst" with formation of *Cenozoic* type of echinofauna are to be observed. Therefore in the development of Postearlyercretaceous echinofauna we can distringuish two stages: Late Cretaceous with formation of Upper Cretaceous type and *Cenozoic* with *Cenozoic* type formation. It is also distinguished the third, Danian/Paleocene, the so-called transition period with "transitional type" of echinofauna [1]. By our point of view separation of transition period with transitional type is not acceptable. We think that Danian-Paleocene complex of echinofauna (especially Danian) is connected very closely to the Upper Cretaceous complex phylogenetically which is confirmed very well through the Caucasian example and thus, it is compound part of the latter one. We think that Senoman-Paleocene echinofauna (more exactly Turon-Paleocene) should be united in the Upper Cretaceous type, while in the *Cenozoic* type - Eocene-Recent one.

As to Danian echinofauna with its development feature so closely connected to the Late Cretaceous echinofauna with the help of echinoids to draw the

Cretaceous/Paleogene boundary between Maastrichtian and Danian we do not consider acceptable.

If we have to begin new geological epoch basing on the beginning of new stage in bios development, the boundary between Cretaceous and Paleogene by the echinoids should be drawn not between Maastrichtian and Danian, but more higher - between Danian and Paleocene or perhaps between Paleocene and Eocene.

I.Javakhishvili Tbilisi State University

REFERENCES

1. *M.Moskvin, A.Solovyev, L.Endelman.* Evolution and Change of Invertebrata and the Mesozoic-Ceinozoic Boundary. M., 1980, 116-167.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

MINERALOGY

G.Nasidze

Minal Composition of Pyroxenes and Amphiboles of the Adjara-Trialeti and their Distribution

Presented by Corr. Member of the Academy G.Zaridze, August 12, 1996

ABSTRACT. **Dependence of chemical composition of volcanogenic rocks, conditions of mineral formation process and influence of geodynamic regime on distribution of minals in minerals of the region has been considered basing on evaluation in minals of pyroxenes and amphiboles of Eocene volcanogenic rocks of Adjara-Trialeti.**

Minal composition of the amphiboles and pyroxenes from Eocene volcanogenic rocks of Adjara-Trialeti was calculated using the method developed by [1,2]. It is established that diopsidic minal in pyroxenes predominates over others, which in the eastern and central parts of the region and in Adjara are distributed in the following succession: hedenbergite, enstatite, Chermak molecule and egerine. In Guria and north-western parts of the region above mentioned succession is preserved with some differences: places of Chermak molecule and egerine are changed.

Chemical composition of volcanics has influence on the distribution of minerals. For example, for volcanics of the region central part enriched by magnesium, content of diopsidic and enstatitic minals is increased. In the places where volcanics are impoverished by Si (NW part of the region) Chermak molecule presents in great quantity. The relation between ferrousity of pyroxenes and quantity of hedenbergite molecule is also recorded. For example in the most ferrous pyroxenes (33.04%) in lavas from the eastern part of the region, quantity of hedenbergite reaches its maximum - (21.42%). Reversed situation is observed in samples with little Fe, e.g. in pyroxenes from several volcanic rocks of Atskuri suite, Adjara and Guria. Genesis of volcanic rocks also has influence on mineral composition. In the Middle Eocenean breccias and their veins, as well as in lavas of the region eastern part, tendency of increasing of egerinic and partially of Chemark molecules is observed.

Quantity of diopside is not always increased. For example in pyroxenes of the region eastern parts, quantity of diopsidic minal fluctuates within 33-45%, whereas in several parts of the region its average composition is 60%. The most poor with diopside minal are pyroxenes from the Paleocene - Lower Eocene sills - 29.75%, but sometimes quantity of enstatitic minal increases - 26.16%. This event testifies redistribution of magnesium between main solid phases, specifically between clyno- and orthopyroxenes.

Condition of mineral formation, and particularly oxydizing potential determining degree of oxydation of Fe and accordingly type of minal and widespread izomorphic replacements has essential influence on minal composition and its quantity in pyroxenes. In the area of maximal extension of rift structure (western part of the region) where partial pressure of oxygen is high, Chermak molecule increases and hedenbergite decreases. Ferric iron, which is formed in the result of oxydation of iron, isomorphically substitutes Al, which represents one of the main constituent of Chermak molecule. But in the case of hedenbergite existence, together with Chemark molecule egerine arises. Negative correlation of hedenbergite and diopside is conditioned by isomorphism.

Up the section quantity of egerine, Chemark molecule and diopside increases, whereas the quantity of hedenbergite and clynoenstatite-decreases.

In the eastern and north-eastern parts of the region varieties of orthopyroxenes were recorded. Great number of bipyroxenes is also found: in the east orthorombic pyroxenes with clynopyroxenic phases and in the north-western parts clynopyroxenes with admixture of orthorombic phase are predominant. In the bipyroxenic varieties of clynopyroxenic phase are represented by enstatite-diopside, rarely by augite. Truly clynopyroxenic varieties are often represented by endiopside.

In the acid effusive rocks of the region eastern part (represented by andesites, andesite-dacites, andesite-basalts and basalts), in comparison with effusive rocks of north-western part (mainly calc-alkaline and alkaline basalts), content of egerinic, hedenbergitic and orthorombic pyroxenes is high. On the other hand, in the latter the main ingredient of pyroxenes - diopside (endiopside) is >50%. For the outskirts of the region approximately equal quantity of Chermak molecule and its steady distribution in ore-type pyroxenes is characteristic.

For the calc-alkaline andesites and andesite-dacites of the central parts of region augites with orthorombic phase, and for the subalkaline basalts-enstatite-diopside and enstatite-diopsides with orthorombic admixture, for andesites and delenites diopsides with orthorombic pyroxenic admixture are characteristic. For the rocks of the Adigeni alkaline series presence of augite is mainly characteristic. In general minal composition of pyroxenes of central parts corresponds more to minal composition of pyroxenes from eastern part, than from north-western part. Only diopside minal is an exception.

By their chemical composition volcanism of Adjara and Guria differs from each other. In Adjara we have subalkaline and calcareous high-aluminous basalts, trachybasalts, trachyandesites, dellenites and andesites, and in Guria - rocks of calcareous, subalkaline and alkaline basaltic andesites, and basaltic-andesitic composition are widespread. In both cases high distribution of endiopsides is characterized rarely - that of endiopsides with admixture of orthorombic phase and augites with orthorombic phase. In the Adigeni suite along with the augites, endiopsid with admixture of orthorombic phase and rarely diopsides and diopsides with orthorombic phase are present. In the Upper Adigeni suite clynopyroxens as well as orthorombic pyroxenes are present. In Guria minal of diopside and Chermak molecule predomanite.

Amphiboles differ from pyroxenes by poor variety of minal composition, occuring in all types of volcanogenic rock of the central parts of the region and are represented by the Fe-hastingsite-pargasite (43 samples) varieties. Fe-edenite-edenite varieties (25 samples) are related to the lower parts of the Paleogenic volcanics. Besides above mentioned, hornblende is one of the main minals. Phases of actinolite, anthophyllite, pyroxenes and rarely tremolite and jadeite also can be found. Amphiboles are mainly represented by pargasite and edinite, and their ferrous varieties are subordinated.

Basing on correlation analyses it was established that in the veins of the middle Eocene breccias of the central part of the region Fe shows positive correlation with oxydizing coefficient, confirming the relation of mentioned minals with oxydizing potential. Pargasitic hornblende can be found in crystal parts of the lava flows, which is also stipulated by high partial pressure of oxygen.

Quantity of hornblende in the young formation is high and reaches 65% in the middle Eocene veins, comparatively to 17.6% in the middle Eocene sandstones. Hornblendes are absent or very rare in lavas of middle Eocene and upper Eocene volcanics are enriched by hornblende (35-65%).

In the lava flows of the eastern region, and in the middle Eocene Fe-hastingsite-pargasrite veins as well as Fe-edinitic varieties can be found. Distribution of hastingsite is even in several types of rocks, and quantities of pargasites varies: in lava flows its quantity is very low and they are enriched by hornblende.

Amphiboles of Adjara are represented by Fe-endenite-endenite varieties, with domination of hornblende reaching 75% in intrusions.

This material allows to conclude that distribution of pyroxenes and amphiboles in the volcanic rocks of the region depends on chemical composition of the melt as well as on thermodynamic conditions of mineral formation and dynamic regime of the formation of the region.

A.Janelidze Geological Institute
Georgian Academy of Sciences

REFERENCES

1. *N.J.Nasidze, R.A.Akhvlediani.* Bull. AS GSSR, **83**, 2, 1976, 417-420 (Russian).
2. *G.J.Nasidze.* Bull AS GSSR, **92**, 2, 1978, 269-372 (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

HYDRAULIC ENGINEERING

J.Kilasonia, G.Khelidze

Topical Problems in Predicting Slope Deformations

Presented by Academician Ts.Mirtskhulava, August 6, 1996

ABSTRACT. **The critical analysis of the methods of estimation of artificial and natural slopes stability is given, as well as short description of the computer model developed by the authors for predicting the deformations of mountain slopes adjoining reservoirs. The ways of the model further development and prospects of its application for the solution of topical problems are scheduled.**

Stability analysis of artificial and natural slopes plays a key role in civil and transport engineering. It assumes special economic and social importance under conditions of constructing and operating hydro-stations. The reliable estimation of stability of the reservoir adjoining mountain slopes subjected to intensive seepage, rheologic processes and seismic loads is of vital importance.

The slope motion can be divided into five groups: collapsing, tipping over, slipping landslide, constrained landslide and flow. The sixth group is - composite movements comprising of two or several above mentioned types combinations. It is very important to determine type of slope motion as it helps to come up with the method of slope stability analysis as well as the reconstruction arrangements. Therefore it is expedient to combine the following two approaches to this problem. The first approach is when a geologist regards the slope motion as a natural process, reasons of its origination, its development and at last the schematization of the slope by means of the mechanical and structural models reflecting the main geological factors: the structure of slope, composition of rocks, exogenous effects and the spatial appearance of these features. The second approach is when an engineer basing on laws of soil mechanics develops the slope stability estimation methods.

The great majority of the slope stability estimation methods are based on the concept of the limit plasticity equilibrium and almost all of them use static approach. Most of slope stability problems, except the simplest ones are statically indeterminable and different assumptions are made in order to provide uniqueness of the solution. It is just due to the specificity of such assumptions that a variety of methods are applied.

In practice methods of Yanbu, Morgenstern-Price, Spencer, Mozhevitinov-Shintemirov are widely used. These methods rely on splitting the collapse prism into interacting hypothetical plane elements. They assume that the dependence between shear and normal forces exerted on the section of the collapse prism is known in advance. This enables to deal with statically determinate problem and thus simplify its solution. The use of the mentioned methods is reasonable for the calculation of stability of slopes constituted of soil or strongly cracked rocks only, as they do not account for contact structural softening (main cracks) in the collapse, or account for it integrally i.e. they regard the rock body as "quasi-solid". The latter is utilised by Freiberg-Kaufman method [3]. In case the crack is filled with soil the strength characteristics of which abruptly differ from the corresponding characteristics of the uncracked part, the combining of this method with that of V.Bukhartsev [4], considering the effect of above mentioned factor on the value of the angle between the interacting forces of the different prism parts, one can arrive at very useful results.

The above mentioned methods of calculation of slope stability are classified as simple engineering methods. They ensure reliable results if the sliding surface is known in advance and the neglecting of time effects is possible at the same time. The most general and complicated cases should be resolved by determining the stress-strain response of the slope with respect to all the exerted forces and the (time) variation of soil (and rock) characteristics, and eventually performing analytical search of slide surfaces and making estimates of thus determined collapse prism's stability. This category of methods has been applied to the earth dams and other artificial slopes stability calculations (V.Lombardo, Yu.Zaretsky). As for the stochastic aspect, it is considered only within the framework of random analysis theory.

The methods and software package developed in the Georgian Scientific Research Institute of Power Engineering under the guidance of J.Kilasonia (co-authors: G.Khelidze, N.Albutashvili) allow to tackle the wider range of problems dealing with stress-strain response and stability of slopes. The methods are mainly aimed on predicting large-scale deformations of the mountain slopes adjoining the mountain reservoirs.

Our software package is based on the recent achievements of the creep theory, fracture mechanics and computer graphics. The finite element method has been used as a main mathematical apparatus underlying its algorithms. The strength, deformation and other physical and mechanical characteristics of the rock materials, along with the corresponding rheologic models' information and packages of seismograms of the different magnitude earthquakes as graphs and tables are the constituent parts of the model too. The software package allows to search and determine potential slide surfaces basing on the numerical analysis of the slope stress-strain state (within the framework of two-dimensional problem) and the minimal geological information taking into account large-scale creep and seismic loads. The seismic loads in the package are considered within wave dynamics numerical scheme and a set of different magnitude seismograms for the reservoir's different operation periods is applied. A computer program builds up the finite element grids for the calculation areas (geological sections) and sends the graphical representation of the interim and final results to the screen or printer. The computer program also carries out the search of slide micro-surfaces and hypothetical cylindrical slide surfaces (the latter in the case of weak rocks). As for the polygonal slide surfaces (in case of rocks), they are fixed manually in the interim stages of calculation by tracing the slide micro-surfaces through finite elements and then calculation is resumed to obtain estimates of the stability of possible collapse prism. On this final stage the above mentioned "engineering methods" can also be applied.

The rheologic models (Burgers, Bingam-Shvedov, Glushko's first, second and third models) being appropriate for the rock materials usually met in the locations of mountain reservoirs are applied having drawn up the table of their parameters' numerical values.

Aimed at expansion and perfection of the software package the following is scheduled:

The adjustment of the slopes' rheologic models parameters.

The computer program for calculation of three-dimensional stress-strain response of hazardous slope sections under seismic loads is to be developed and included in the package.

The three-component seismograms' package shall be expanded.

The program of modelling of the seepage processes will be essentially altered.

The computer simulation of the well-known slope collapse histories will be carried out and the results used for the adjustment of algorithms.

The engineering provisions for the ensuring of slopes' stability will be developed and the computer simulation of collapse histories will be carried out in order to check the efficiency of above provisions.

The methods of waterside slopes safety factors estimation for reservoirs' different operation stages based on the probabilistic reliability theory will be elaborated.

Georgian Scientific Research Institute of Power Engineering

REFERENCES

1. *I.J.Juan.* Ustoychivost Zemlyanikh Otkosov. M., 1988, 240.
2. A.*L.Mozhevitinov, M.Shintemirov.* Izvestia VNIIG im. B.E.Vedeneeva, **9**, 1970, 11-22.
3. E.*A.Freiberg, M.D.Kaufman.* Izvestia VNIIG im.Vedeneeva, **147**, 1981, 114-123.
4. B.*N.Bukhartsev.* Izvestia VNIIG im. B.E.Vedeneeva, **130**, 1979, 9-13.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

T.Magrakvelidze, N.Bantsadze, N.Lekveishvili

Influence of Artificial Roughness on Heat Transfer to Turbulent Mixed Liquid in a Pool

Presented by Academician M.Salukvadze, July, 29, 1996

ABSTRACT. **Heat transfer from thorus shaped smooth and roughened surfaces to mixing water in a pool has been investigated experimentally. The surface was roughened by means of wire rings, fixed at a constant pitch-to-wire diameter ratio of 10/1. Wire diameter was 0.5 mm. It was shown that artificial roughness increased heat transfer intensity about 1.6 times.**

There is widely used apparatus with mixers in modern technological processes, and particularly in chemical engineering. The efficiency of such kind of apparatus considerably depends on heat transfer intensity. From this point of view the intensification of heat transfer, when liquid is turbulently mixed in a pool, is of great practical interest.

As it is well-known from references, by applying the method of artificial roughness in canals, heat transfer coefficient of turbulent flow was increased about 2.5 times [1]. This method is effective also in the case of subcooled liquid boiling in canals [2]. At the same time, the influence of artificial roughness on heat transfer to turbulent mixed liquid in a pool has not been investigated.

The aim of the present investigation is to confirm experimentally influence of artificial roughness on heat thansfer of mixed turbulent liquid flow.

To realize this an experimental apparatus was created (Fig. 1). The main part of the apparatus is cylindrical vessel 1, made of stainless steel. The diameter of the vessel was 200 mm, the height-300 mm. The vessel was equipped with stainless steel impeller mixer 2. There was used a flat-paddle mixer with two paddles. The paddles were 10 mm in width. The diameters of the mixer were 100 and 120 mm. The mixer was connected with electric motor shaft. Electric motor 3, mounted on the cover of the vessel,was rotating the mixer.

Thorus shaped heating tube 4 was made of stainless steel. The outside diameter of the tube was 10 mm, wall thickness-1 mm. The average diameter of thorus was 140 mm. Heating tube was located coaxially with mixer's shaft in horizontal position at a distance of 50 mm from the bottom of the vessel. At the ends of the heating tube there were soldered copper conductors 5. Cylindrical vessel was filled with distilled water to the level of 250 mm. To keep water temperature constant and to maintain stationary process, the side wall of the vessel was cooled by means of water jacket 6. The temperature of the water in the vessel was measured by using chromel-alumel thermocouple placed in a stainless steel pocket 9 filled with transformer oil.

The test section was heated by means of low voltage AC from OCY-20 transformer. The ratio of the heating power was controlled by PHO-250-10 autotransformer.

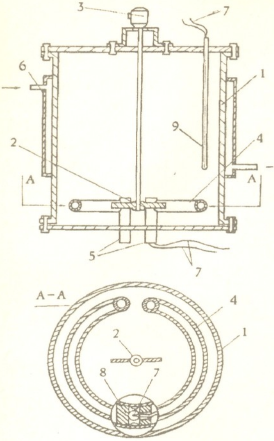


Fig. 1. Experimental apparatus
1 - cylindrical vessel; 2 - impeller mixer; 3 - motor; 4 - heating tube; 5 - conductors; 6 - cooling jacket; 7 - thermocouples; 8 - teflon chamber; 9 - thermocouple pocket.

Low voltage AC electricity was supplied to the heater tube from the OCY-20 transformer. The ratio of the heating power was controlled using PHO-250-10 autotransformer.

Rotational velocity of the electric motor was controlled by means of LATR-250-2 autotransformer.

Cylindrical vessel was equipped with a water level indicator.

The strength of electric current feeding heated tube was measured by B7-37 electronic multimeter, connected to the current transformer YTT-6. A voltage drop across the heating tube was measured by B7-21A electronic voltmeter.

Mixer rotating electric motor voltage was measured by B7-16A, current-with B7-35 electronic multimeters.

Stroboscopic tachometer was used to measure mixer's r.p.m.

The temperature was measured with Chromel-Alumel thermocouples.

The measurments of the thermocouple e.m.f. were made by B7-21A electronic millivoltmeter.

To determine heat transfer coefficient, temperature of heating tube inner wall was measured at three different positions. Thermocouples were fixed in teflon chambers sealed properly in the heating tube (Fig.1, pos. 7,8). Chambers were necessary to avoid air convection in the inner space of the heating tube.

Outside the wall temperature of the heating tube was calculated according [3].

Liquid bulk temperature in the vessel was also measured using chromel-alumele thermocouple.

Heat transfer coefficient was calculated from the equation

$$\alpha=\frac{Q}{F(T_w-T_f)},$$

where $Q=I\cdot\Delta U$-is a quantity of heat emitted in the heating tube, kJ; I-is the strength of the current in the heating tube, A; ΔU-voltage drop across heating tube, V; F-outside wall surface area of the heating tube, $F=13.5\cdot10^{-3}$ m^2; T_w - outside wall temperature of the heating tube, K; T_f - liquid bulk temperature in the vessel, K;

Experiments were carried out both for smooth and for roughened tubes. Tube was roughened by fixing wire rings on it. Wire diameter was 0.5 mm, pitch-to-wire diameter ratio $s/d=10$.

Experimental results for smooth and roughened surfaces are plotted in Fig. 2 as a relation $A=f(Re_c)$, where

$$A=Nu\cdot Pr^{-0.33}\left(\frac{D}{d_m}\right)^{-0.25}\left(\frac{D}{H}\right)^{-0.25}\left(\frac{\mu}{\mu_w}\right)^{-0.14}$$

here $Nu=\alpha\cdot D/\lambda$ is Nusselt number; $Re_c=n\cdot d_m^2/\nu$- Reynolds number; $Pr=\nu/a$-Prandtl number; D is a diameter of cylindrical vessel, *m*; d_m is diameter of mixer, *m*; n is a rotational frequency of the mixer 1/*s*; λ-is thermal conductivity coefficient of heating tube wall, *W/m·K*; a is a thermal diffusivity of the liquid, m^2/s; H is a level of liquid in the vessel, *m*; ν - kinematic viscosity of liquid m^2/s; μ is a dynamic viscosity of liquid at T_f, kg/m·s; μ_w is a dynamic viscosity of liquid at T_w, kg/m·s;

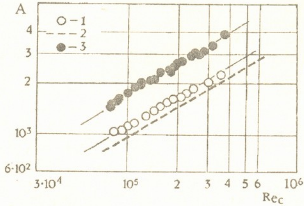


Fig. 2. Relation between heat transfer intensity and the Re_c number.
1 - smooth surface; 2 - smooth surface [4];
3 - roughened surface (d=0.5 mm, s=5mm).

From the diagrams (Fig. 2. pos. 1,2) it can be seen that results of the present investigation for smooth heating tube are in good agreement with the data, obtained by F.H.Chilton et al. [4].

At the same time, heat transfer intensity for roughened tube is about 1.6 times higher, than for smooth one (Fig.2).

Here it must be mentioned, that in spite of considerable increasing of heat transfer intensity, in this case the method of artificial roughness was found less effective than in case of turbulent flow in canals.

This can be explained so: liquid flow in apparatus with mixer is three-dimentional [5] and turbulence can be considered isotropic in the first approximation.

It means that heat transfer coefficient consists of three components:

$$\alpha=\alpha_c+\alpha_r+\alpha_a,$$

here α_t, α_r and α_a are heat transfer coefficients corresponding to tangential, radial and axial components of the liquid flow.

So far as the elements of artificial roughness were positioned perpendicular to the tangential component of the stream, we can make an assumption that artificial roughness makes intensive mainly α_c. The influence of artificial roughness on the other two components α_r and α_a is slight.

Basing on the experimental data, we can make a conclusion that application of the artifical roughness method in described conditions increases heat transfer considerably.

Institute of Control Systems
Georgian Academy of Sciences

REFERENCES

1. *V.I.Gomelauri.* Works of the Institute of Physics of the Academy of Sciences of the Georgian SSR. **9**, 1963, 110-145.
2. *V.I.Gomelauri, T.Sh.Magrakvelidze.* Heat Engineering, **6**, 1976, 4-7.
3. *M.A.Mikheev, I.M.Mikheeva.* Heat Transfer. Moscow, 1973.
4. *F.H. Chilton, T.W.Drew, R.H. Jebens.* Ind. Eng. Chem. **36**, 1944, 510-516.
5. *V.V.Konsetov.* Journal of Engineering Physics. **10**, 2, 1966.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

AUTOMATIC CONTROL AND
COMPUTER ENGINEERING

G.Vachiberidze

On Some Aspects of Digital Processing and Improvement on Recognition Parameters of the Earth Surface Radio Location Images

Presented by Academician M.Salukvadze, August 9, 1996

ABSTRACT. **The problems of digital processing of radio location images are considered conformably to the devices of air-borne computer complexes operating in real time. In order to process running images of the earth surface, the method of median digital filtration is used, its algorithms being implemented on sorting elements of programmable memory device connected as conveyer sorting circuit. It is shown that median filtration with optimum size two-dimensional window enables to increase digital processing quality and recognition accuracy of radio location images.**

Digital processing of radio location images is in most cases stipulated by the solution of number of applied problems. One of them is recognition of the given area on the earth surface basing on the traditional method of recognition theory-comparison of running image with standard one. Usually standard image is created using the method of mathematical modelling on the basis of topographic maps and aerial photographs, while terrain sounding with radar beams is used for getting the current image. The latter method has its definite advantages compared to other ones, such as: high accuracy of image recognition, stability and reliability of the obtained results [1,2,3].

The variation of radio location pictures caused by different jammings and distortions is till today a factor, interfering accurate recognition of radio location images. This certainly determines the technique of radio location image processing and recognition, algorithms and device implementation.

From the above said it follows that the development of air-borne digital device of radio location image recognition is based on simple algorithms constructed on the principle of correlation-extremum systems of comparision [4]. Recognition problem is based on two decisive rules.

The first implies the determination of Hemming distance between running image and each standard and finding of modular function minimum by the formula.

$$F_l = \min_k \sum_i \left| Y_{ik}^M - Y_i^T \right|, \qquad (1)$$

where Y_{ik}^M, Y_i^T are the values of standard matrix and current image elements, respectively; i, k is matrix size.

According to the second decisive rule the maximum of logical comparison function of running and standard image elements is determined by the formula

$$F_2 = \max_k \sum_i Y_{ik}^M \bigcap Y_i^T , \quad (2)$$

where $\bigcap$ is a logical comparison operation satisfying the following conditions

$$Y_{ik}^M \bigcap Y_i^T = \begin{cases} 1, \text{ if } Y_{ik}^M = Y_i^T \\ 0, \text{ if } Y_{ik}^M \neq Y_i^T \end{cases}$$

Image recognition problem can be solved at simultaneous or separate employment of the above presented rules. As a result of operations (1) and (2), such extreme values of K are chosen when functions (1) and (2) will assume minimum and maximum values respectively.

In order to get high recognition accuracy the pre-improvement of running image is done using digital filtration chiefly. In air borne digital systems where running radio location images of an area are generally formed, the comparatively simple, but effective and speedy filtration algorithms are more advantageous. Median filtration of images is referred to such types of processing of running images of the terrain [5,6].

The types of median filters, filtration methods, algorithms and device implementation are considered in [1,2,5,7]. The principle of median filter operation is as follows: in a set of discretized images one- or two-demensional window (aperture) with odd number of images is chosen, the values of all images are substituted with median (mean) values in its center. Then the window slides over the whole set and thus the values of set images are substituted with values of medians. In this process, linear, square and cross-like median filters are mainly used.

In a test sample of radio location coordinator digital computer complex worked out by the authors [8] a conveyer-type median filter device is presented. Its working algorithm provides digit-by-digit search of median using a small set of simple processor elements. The device is charachterized by high filtration frequency giving a good opportunity of image processing in real time. The device is protected by the author's certificate [9]. The so-called systological algorithm worked out here is implemented with conveyer sorting circuit on sorting elements (Fig.1). Each element comprises a programmable memory device, compares values and rearranges two numbers (X and X) in the ordered set of elements, placing the greater of the compared numbers in the right and the smaller ones-in the left according to the following simple algorithm

if $X_1 > X_2$ then $Y_1 = X_2$ and $Y_2 = X_1$;
if $X_1 \leq X_2$ then $Y_1 = X_1$ and $Y_2 = X_2$.

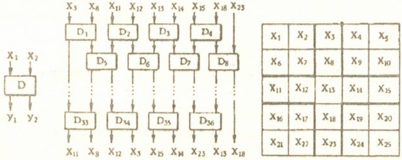


Fig. 1.

Thus, on the output of conveyer sorting scheme an ascending sequence of number values is obtained, its mean element is median. Consequently, in order to draw up a sorting circuit for a cross like aperture of 5x5 two-dimensional windows, 9 X 4 =36 memory elements are needed. The expences for the device as a whole is 4 times less than for the existing one.

In order to show operation principle of the sorting circuit, in the right side of Fig.1, a square window 5 X 5 is designed, in its cross-like aperture image elements of arbitrary values are located. Suppose that X_3=5, X_8=2, X_{11}=1, X_{12}=3, X_{13}=10, X_{14}=7, X_{15}=6, X_{18}=13, X_{23}=9. Sorting circuit arranges the elements in ascending order 1, 2, 3, 5, 6, 7, 9, 10, 13. Central element of this numerical sequence is the fifth element, i.e. X_{15}=6, which is considered as a median of the given aperture.

Experimental investigations of digital processing of the earth surface image showed the advantages of median filtration to such methods as quadratic averaging, hysteresis smoothing, etc. Fig.2 shows the mechanism of operation of video pulse median filtration with windows of different sizes. As it is seen from the figure, median filter provides pulse noise smoothing even in the case of small windows [7].

As a result of theoretical analysis and experimental investigations, the dependence of image brightness dispersion (S) and correct recognition frequency (η) to the size of two-dimensional median filter was determined in order to estimate the efficiency of algorithms of preliminary processing and recognition of radio location images (Fig3).

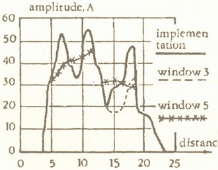


Fig.2

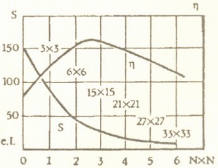


Fig.3

Brightness dispersion of primary and transformed images or standard deviation was calculated by the formula:

$$S = \frac{1}{N}\sum_{i=1}^{N}\left(X_i - \overline{X}_i\right)^2,$$

where X_i is brightness of i-th point, N is the number of image points, $\overline{X}_i = \frac{1}{N}\sum_{i=1}^{N} X_i$ is mathematical expectation or mean value of X. The given figure clearly shows that brightness dispersion is abruptly decreased with the increase of median filter window size causing radio location image information drop and disappearance of detail in the picture.

In image recognition problem the optimum size of median filter is that one giving the highest frequency of correct recognition η in a series of tests. In our case correct recognition frequency is determined with Hemming distance, the minimum of which is taken as the sign of running radio location image proximity to the standard one. For the considered concrete case the range from 9 X 9 to 15 X 15 elements is assumed to be optimum size of median filter aperture, where correct recognition frequency reaches the maximum (Fig.3).

Theoretical analysis, bench and full-scale tests, experimental investigations carried out on stationary digital computer complex of radio location station enable to make the following solutions:

Radio location image employment for recognition purposes needs its indispensable preliminary processing.

Median filter is assumed to be and effective method of radio location image processing.

Median filtration with optimum size two-dimensional window enables to create a device operating in real time and to increase the accuracy of radio location image recognition.

Hemming distance minimum used in the decisive recognation rule makes possible to achieve on flying vehicles the necessary accuracy of image in different climate, weather, season and other conditions.

Scientific Research Institute of Automatics and
Computing Technics

REFERENCES

1. *G.D. Vachiberidze, L.V.Petrov, A.A.Nemsadze.* Defence Technology, **2**, 1987, 91-92.
2. *G.D.Vachiberidze, L.V.Petrov, A.A.Nemsadze.* Problems of Defence Technology, **6**, 1987, 30-31.
3. *V.A.Kovalevski.* Methods of Optimum Solutions in Image Recognition. M., 1976, 221
4. *G.D.Vachibelidze, A.L. Chkhenkeli, L.V.Petrov, et al.* Scientific Technical Report, M., 1982, 9-11.
5. *U.Pratt.* Digital Image Processing. 2, 1982, 480.
6. *G.D.Vachiberidze, Yu. I. Mshvenieradze et al.* Scientific Technical Report, Tbilisi, 1989, 32-34.
7. *G.D.Vachiberidze, L.V.Petrov et al.* Scientific Technical Report, Tbilisi, 1988, 25-37.
8. *G.D.Vachiberidze, L.V.Petrov et al.* Scientific Technical Report, Tbilisi, 1988, 25-29.
9. *G.D.Vachibelidze, L.V.Petrov et al.* Object Image Processing Device. A.C. No 1495830, 1987.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3 1996

AUTOMATIC CONTROL AND
COMPUTER ENGINEERING

M.Karkashadze

Stadial Indices of the Parallel Type CS Reliability and Efficiency

Presented by Academician V.Chichinadze, September 2, 1996

ABSTRACT. **In the work there are established the fundamental indices of the redundant joint usage engineering systems with account of the reliability in the stadial mode. The system is considered as the system of the mixed type mass service (with losses and expectations), with a united queue, common resourcess and group service.**

In the work [1] the indices of reliability and efficiency of multiprocess and multimachine computer (CS) in the Laplace transformation mode were obtained. The aim of the present work is to determine these indices in the stadial mode being themselves its continuation.

Designating:

$$R_0^{(k)} = \lim_{s\to 0} s\overline{R}_0^{(k)}(s);\quad R_i = \lim_{s\to 0} s\overline{R}_i(s);\quad P_{ij}^{(k)}(0) = \lim_{s\to 0} s p_{ij}^{(k)}(s, 0);$$

$$F_i^{(k)} = \lim_{s\to 0} s F_i^{(k)}(s) =$$

$$= \sum_{\nu=1}^{n}\sum_{\varepsilon=1}^{\eta_\nu}\sum_{\gamma=\varepsilon}^{c+\varepsilon} P_{\nu\eta_{\nu\gamma}}^{(\gamma)}(0)\left[A_k^{(c)}\overline{h}_{\nu i}^{(\varepsilon)}(0) + \sum_{r=1}^{c} B_{\gamma-\varepsilon,k}^{(c)}(s_r)\overline{h}_{\nu i}^{(\varepsilon)}(-s_r r)\right]\delta(\eta_{\nu\gamma} = \varepsilon); \tag{1}$$

from the equation systems of the non-stadial mode (demonstrated in [1]) we move to the equation systems of the stadial mode (these equations can be obtained directly):

$$[\mu + (1 - \delta_{oc})\lambda]R_0 - (1 - \delta_{oc})\tau R_0^{(1)} = a_1 R_1; \tag{2}$$

$$-\lambda R_0^{(k-1)} + [\mu + (1 - \delta_{kc})\lambda + k\tau]R_0^{(k)} - (1 - \delta_{kc})(k + 1)\tau R_0^{(k+1)} = 0,\ k = \overline{1,c}; \tag{3}$$

$$-\mu R_{i-1} + [\lambda + (1 - \delta_{in})\mu + a_i]R_i - (1 - \delta_{in})a_{i+1}R_{i+1} = F_i^0,\ i = \overline{1,n}; \tag{4}$$

$$P_{i\eta_{ik}}^{(k)}(0) - F_i^{(k)} - \delta_{k1}\lambda R_i - \delta_{i1}\mu R_0^{(k)} = 0,\ i = \overline{1,n},\ R_0^{(0)} = R_0 \tag{5}$$

$$R_0^{(k)} = 0,\ k > c;\ P_{i\eta_{ik}}^{(k)}(0) = 0\ (k > c + \eta_{ik}).$$

In (1), (2), (3) and (4) all the designations given in [1] are preserved. We're going to repeat some of them: m - the quantity of the parallely working processing equipments (PE), n-m - the quantity of the reserve processing equipments (RE), λ - the intensity of the stream of demands (orders), α_1- the intensity of the PE failures, α_2- the intensity of the RE failures, μ- the intensity of the restoration of the failed equipments, $h_{ij}^{(k)}(v)dv$ - the probability of serving the demands to be completed within the time interval (v, v+dv) by the service system (SS) being in the state j, under condition that initially it was in the state i; R_i - the probability of the fact that in the stadial mode the quantity of the equipments in good working order equals i and the orders are absent. $R_0^{(k)}$ - the probability of disrepair of all the equipments in the stadial mode, when K orders are received within the restoration time; $p_{ic}^{(k)}(u)du$ - the quantity of the orders in the stadial mode equals K_i among them l ($l \leq \min(i, k, m)$) is served during the time u, however their service has started since the moment b - time $u = 0$, being in the state i $(i = \overline{1,n})$.

Summing up the left and the right sides of equations (2), (3), (4) and (5) we get sure that they are linearly dependent. We use the equation of norming as one missing equation.

$$\sum_{i=1}^{n} R_i + \sum_{k=0}^{c} R_0^{(k)} + \sum_{i=1}^{n}\sum_{l=1}^{\eta_i}\sum_{k=1}^{c+1} P_{il}^{(k)} = 1 \qquad (6)$$

where

$$P_{il}^{(k)} = \lim_{s\to 0} s\overline{P}_{il}^{(k)}(s) = \sum_{\gamma=1}^{c+1} \mathrm{P}_{i\eta_{i\gamma}}^{(\gamma)}(0)\left[T_i^{(l)} A_{k-l}^{(c)} + \sum_{r=1}^{c} B_{\gamma-1,k-1}^{(c)}(s_r)\left(\frac{1-\overline{h}_i^{(l)}(-s_r\tau)}{-s_r\tau}\right)\right], \qquad (7)$$

$$T_i^{(l)} = -\left(\overline{h}_i^{(l)}(s)\right)' \; s = 0$$

The expression of the other parts of the above demonstrated formulae of quantity is presented in [1].

Our further problem is to present the solution of equation systems (2)-(6) in a more convenient form for practical usage and in the form of easily realizable algorithms of computer (their reduction to one equation system, $p^{(\alpha)}(0)$ in parcticular.

In the beginning we solve equation system (2) and (3), $R_0^{(k)}(k = \overline{0,c})$ in particular. We take matrix $(\widetilde{D})$, the coefficients of the unknowns being a three-diagonal square matrix of rank $c + 1$:

$$\widetilde{D} = \begin{bmatrix} \mu+(1-\delta_{0c})\lambda & -(1-\delta_{0c})\tau & 0 & \cdots & 0 & 0 \\ -\lambda & \mu+(1-\delta_{1c})\lambda+\tau & -2(1-\delta_{1c})\tau & \cdots & 0 & 0 \\ 0 & -\lambda & \mu+(1-\delta_{2c})\lambda+2\tau & \cdots & 0 & 0 \\ \cdots & \cdots & \cdots & \cdots & \cdots & \cdots \\ 0 & 0 & 0 & \cdots & -\lambda & \mu+c\tau \end{bmatrix}$$

Solution (2) and (3) can be obtained through the determinants of the submatrices of matrix $\widetilde{D}$ obtained by two ways: on moving from up to down $(\widetilde{D}_n)$ and on moving upwards from bottom $(\widetilde{\Delta}_n)$. The determinants are in the initial conditions and represent the following recurrent correlation:

$$\widetilde{D}_0 = 1;\ \widetilde{D}_{-1} = 0,\ \widetilde{D}_1 = \mu + (1-\delta_{c0})\lambda,\ \widetilde{D}_{n+1} = b_{n+1}\widetilde{D}_n - \widetilde{a}_{n+1} d_n \widetilde{D}_{n-1};$$

$$\widetilde{\Delta}_0 = 1;\ \widetilde{\Delta}_{-1} = 0,\ \widetilde{\Delta}_1 = \mu + cr,\ \widetilde{\Delta}_{\underline{n+1}} = b_{c-n+1}\widetilde{\Delta}_n - a_{c-n+2} d_{c-n+1}\widetilde{\Delta}_{n-1};$$

$$n = \overline{0,c}.$$

where

$$\widetilde{a}_1 = 0,\ \widetilde{a}_2 = \widetilde{a}_3 = \ldots = \widetilde{a}_{c+1} = -\lambda,\ d_0 = 0,$$

$$b_{n+1} = \mu + (1-\delta_{nc})\lambda + n\tau,\ d_{n+1} = -(1-\delta_{nc})(n+1)\tau,$$

$$b_c = \mu + c\tau,\ d_{c+1} = 0.$$

Determinant $|\widetilde{D}_{c+1}| = b_{c+1}\widetilde{D}_c - a_{c+1}d_c\widetilde{D}_{c-1}$.

solutions (2) and (3) get the form:

$$R_0^k = \lambda^k a_l R_l \tilde{\Delta}_{c-k} |\tilde{D}_{c+l}|, \ k = \overline{0,c}.$$

Afterwards we consider the solution of system (4), R_i $(i = \overline{1,n})$ in particular, substituting the value of R_0 presenting it in such a way.

$$\left[\frac{\lambda + (1-\delta_{ln})\mu + a_l - a_{l\mu}\tilde{\Delta}_c}{|\tilde{D}_{c+l}|}\right] R_l - (1-\delta_{ln})a_2 R_2 = F_l^{(0)}$$

$$-\mu R_{i-l} + [\lambda + (1-\delta_{in})\mu + a_i]R_i - (1-\delta_{in})a_{i+l}R_{i+l} = F_i^{(0)}, \ i = \overline{2,n} \qquad (8)$$

We designate by $D^{(0)}$ three diagonal square matrix of n - rank coefficients, with the unknown R_i and by $|D_m^{(0)}|$ - its determinant.

$$D^{(0)} = \begin{bmatrix} b_l & c_l & 0 & 0 & 0 & \cdots & 0 & 0 \\ -\mu & b_2 & c_2 & 0 & 0 & \cdots & 0 & 0 \\ 0 & -\mu & b_3 & c_3 & 0 & \cdots & 0 & 0 \\ \cdots & \cdots & \cdots & \cdots & \cdots & \cdots & \cdots & \cdots \\ 0 & 0 & 0 & 0 & 0 & \cdots & -\mu & b_n \end{bmatrix}$$

Here

$$b_l = \lambda + (1-\delta_{ln})\mu + a_l - a_l\mu\tilde{\Delta}_c / |\tilde{D}_{c+l}|,$$

$$C_\eta = -(1-\delta_{\eta n})a_{\eta+l}, \ \eta = \overline{1,n}; \ b_\eta = \lambda + (1-\delta_{\eta n})\mu + a_\eta, \ \eta = \overline{2,n}$$

Analogically the solution of equation system 4 can be obtained through the determinants of the submatrices of matrix $D^{(0)}$, obtained by two ways: on moving from up to down $(D_m^{(0)})$ and upwards from bottom (Δ_m).

It gets the following form:

$$R_i = \left[\Delta_{n-i}^{(0)} \sum_{\sigma=1}^{i} F_\sigma^{(0)} D_{\sigma-1}^{(0)} \mu^{i-\sigma} + D_{i-1}^{(0)} \sum_{\sigma=i+1}^{n} F_\sigma^{(0)} \Delta_{n-\sigma}^{(0)} \prod_{\eta=i}^{\sigma-1}(-C\eta)\right] / |D_n^{(0)}|; \qquad (9)$$

Here

$$D_0^{(0)} = 1, \ D_l^{(0)} = b_l,$$

$$D_m^{(0)} = \left[\left(\frac{-\delta_{lm} a_l \mu \tilde{\Delta}_c}{|\tilde{D}_{c+l}|}\right) + \lambda + (1-\delta_{mn})\mu + a_m\right] D_{m-l}^{(0)} - \mu(1-\delta_{m-l,n})a_m D_{m-2}^{(0)};$$

$$\Delta_0^{(0)} = 1, \ \Delta_l^{(0)} = \lambda + a_n,$$

$$\Delta_m^{(0)} = \left[-\delta_{mn}\left(\frac{a_l \mu \tilde{\Delta}_c}{|\tilde{D}_{c+l}^{(0)}|}\right) + \lambda + (1-\delta_{n-m+l,n})\mu + a_{n-m+l}\right] \Delta_{m-l}^{(0)} -$$

$$-\mu(1-\delta_{n-m+l,n})a_{n-m+2}\Delta_{m-2}^{(0)};$$

$$|\Delta_n^{(0)}| = [(-\delta_{l,n}a_l\mu\tilde{\Delta}_c / |\tilde{D}_{c+l}|) + \lambda + a_n] D_{n-l}^{(0)} - \mu a_n D_{n-2}^{(0)}; \ m = \overline{1,n}.$$

After substituting $R_0^{(K)}(k = 1,c)$, $F_i^{(k)}(K = 0,c)$ and R_i $(i = 0,n)$ (5) and (6) we get the system of Linear algebraic equations, $p_{i\eta_{ik}}^{(k)}(0)$ in particular, which is soluted by the usual method, that's why we are not demonstrating it here.

Kutaisi Technical University

REFERENCES

1. *I.Mikadze, M.Karkashadze*. Bull. Georg. Acad. Sci. (to appear).
2. *Dzh. Riordan*. Veroyatnostnie Sistemy Obsluzhivanya M., 1966. (Russian).
3. *R.Kakubava*. Thesis for Doctor Degree. Tbilisi, 1994. (Russian).
4. *B.Gnedenko*, *I.Kovalenko*. Vvedenie v Teoryu Massovogo Obsluzhivanya. M. 1987. (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

AUTOMATIC CONTROL AND
COMPUTER ENGINEERING

I.Mikadze, L.Mirtskulava

On the Realizability of Message Transmission through Unreliable Data-Transmission Channel

Presented by Academician V.Chavchanidze, October 1, 1996

ABSTRACT. **In the present article we investigate mathematical model of data-transmission channel (DTC), subjected to distortions (failures). The time between neighbouring failures is distributed according to Erlang ratio. The method of enhance of reliability and efficiency of message transmission through unreliable DTC is suggested.**

In the articles [4-7] there were considered analytical model of realizability of task operation by the technical system (queueing system - QS) in a given time with account of possibility of failures inception which are distributed according to Poisson ratio. It is well known that under arbitrary distribution ratio of time between neighbouring failures the most better approach to real system is provided by the creation of the most adequate realistic probability model. However the creation of analytical methods of investigation of such generalized mathematical models is connected with great difficulties. One of the most effective methods is approximation of empiric (sometimes theoretical) distribution ratios of random variables according to Erlang ratio. In practice two basic numerical characteristics of the distribution function (DF) are generally considered: expected value and variance. It is well known that the approximation of the DF with a given meanings of foregoing characteristics may be according to Erlang ratio if the coefficient of variation is less than 1 or otherwise to hyperexponential distribution.

In a given article there is investigated the mathematical model of QS which is presented in the form of DTC where times of neighbouring failures are distributed according to Erlang ratio.

Let messages with fixed length T are transmissed through DTC subjected to distortions (failures),with the purpose of optimum organization of transmission process (in minimum possible time) a message as a rule is broken up into separate blocks; each block is supplied with control information for any failures (distortions) detection being in link channel; the checking of correct transmission is made at the end of reception of each block; if distortion there exists repeated transmission untill the block will not be received without distortions, but no more than determined times amount after a channel is transmissed for repair after which information transmission is renewed from distorted block.

The purpose of this article is to determine DF of random length message transmission time $\Phi(t)$ in independent of quantity of blocks of message and quantity of repeated transmissions under the given characteristics of DTC. We denote:

T - length of message; n - quantity of blocks in message;

l - quantity of phases under Erlang distribution of failures, i.e. the scheme of failures origin takes place according to which scheme process of failure can pass l phases (stages), before they will really origin; $F(u) = 1(t-\tau_b)$–DF of block length in common with control categories ($\tau_b = T/n$ – block length): r - permissible quantity of block transmission replications, after which DTC is transferred to repair; $G(u)$ - DF of recovery time; α - a traffic rate of each phase distribution i.e. the duration of time intervals between successive moments of distortions origin have Erlang distribution ($A(u)$):

$$A(u) = \frac{\alpha(\alpha u)^{l-1} e^{-\alpha u}}{(l-1)!}$$

For analytical description of the above-mentioned model we introduce: $\Phi_j^{(kv)}(t,x,T)$ - the probability that fixed length T message transmission (which consists of n blocks, each of them has a length τ_b) will be completed beginning with j-th block for a time less than t if : 1) at the moment of $t = 0$ DTC was in k phase on failures and was transmissed x-th ($x \in [0, \tau_b]$) part of j-th block; 2) j-th block transmission was realized to try without distortions ν times.

According to definition:

$$\Phi_j^{(kv)}(t,x,T) = \begin{cases} \Phi_{j+1}^{(kv)}(t,0,T), & x = \tau_b \\ 0, & x > \tau_b \end{cases}$$

$$k = \overline{1,l}, \quad \nu = \overline{1,r}, \quad j = \overline{1,n};$$

We denote: $\Phi(t,T)$–DF of probability of fixed length $T = n\tau_b$ message transmission time and if length of message is arbitrary denoted by with $\widetilde{F}(u)$ DF we denote with $\Phi(t)$. The following relationship takes place:

$$\Phi(t) = \int_0^\infty d\widetilde{F}(u)\Phi(t,u)$$

Were $\Phi(t,u) = \sum_{\nu=1}^{r}\sum_{k=1}^{l} \Phi_l^{(kv)}(t,0)/l$ is DF of fixed length message transmission time.

Later on we'll denote $\Phi_j^{(kv)}(t,x,T)$ without argument T that is through $\Phi_j^{(kv)}(t,x)$.

Above-mentioned model can be described by the following system of integral equations:

$$\overline{F}(x)\Phi_j^{(kv)}(t,x) = \int_0^t e^{-\alpha u}\Phi_{j+1}^{k1}(t-u,0)d_u F(x+u) +$$

$$+\int_0^t \alpha e^{-\alpha u} d_u \overline{F}(x+u)\Phi_j^{(k+1,v)}(t-u,x+u), \tag{1}$$

$$j = \overline{1,n}, \quad k = \overline{1,l-1}, \quad \nu = \overline{1,r};$$

$$\Psi_j^{(lv)}(t,x) = \int_0^t e^{-\alpha u}\Phi_{j+1}^{(l1)}(t-u,0)d_u F(x+u) +$$

$$+\sum_{k=1}^{l}\sum_{p=0}^{\infty}\int_0^t \alpha e^{-\alpha u} du \int_0^{t-u} d_\upsilon F(x+u+\upsilon)\left[B_*^{(pl+k-1)}(\upsilon) - B_*^{(pl+k)}(\upsilon)\right]\Phi_j^{(k,\nu+1)}(t-u-\upsilon,0), \tag{2}$$

$$\nu=\overline{1,r-1}\ ;\ j=\overline{1,n};$$

$$\Psi_j^{(lr)}(t,x)=\int_0^t e^{-\alpha u}\Phi_{j+1}^{(l1)}(t-u,0)d_uF(x+u)+$$

$$+\int_0^t\left(1-e^{-\alpha u}\right)d_uF(x+u)\int_0^{t-u}\Phi_j^{(1,1)}(t-u-\upsilon,0)dG(\upsilon) \qquad (3)$$

$$j=\overline{1,n}$$

Here $B_*^{(k)}(\upsilon)$ - k-multiple convolution $B(\nu)=1-e^{-\alpha\nu}$; Let

$\psi_j^{(k\nu)}(t,x)=\bar{F}(x)\Phi_j^{(k\nu)}(t,x)$; $\bar{F}(x)=1-F(x)$; $F\left(\tau_b^{+}\right)=1$; $F(\tau_b)=0$.

Then

$$\Psi_j^{(k\nu)}(t,x)=\begin{cases}\Phi_j^{(k\nu)}(t,0), & x=0\\ \Phi_{j+1}^{(k\nu)}(t,0), & x=\tau_b\\ 0\ \ , & x>\tau_b\end{cases}$$

$$k=\overline{1,l}.\ \ \nu=\overline{1,r}\ ,\ j=\overline{1,n};$$

The boundary conditions have the following form:

$$\Psi_{n+1}^{(k\nu)}(t,x)=\Phi_{n+1}^{(k\nu)}(t,0)=1.\ k=\overline{1,l},\quad \nu=\overline{1,r},x\in[0,\tau_b]$$

We'll explain one of the equations, e.g. (2) relatively to $\Phi_j^{(k\nu)}(t,x,T)$-completion of message transmission ("task solution") under above -mentioned conditions for a given time. For that we'll present a random event A - the completion of "task solutions" in form of two injoint A_1 and A_2 events. Complicated event A_1 consists in, that:

1) transmission of j-th $\left(j=\overline{1,n}\right)$ will be completed for a time $x+u$ under condition that it has not been completed for a time x (probability of this event equals to $d_uF(x+u)/\bar{F}(x)$); **2)** for a time u should't take place branching of QS to next phase on failures ($exp(-\alpha u)$); **3)** for a time $t-u$ the transmission of the whole remained part of message starting from $j+1$-th block when QS is in l state on failures (probability of this event equal $\Phi_{j+1}^{(l1)}(t-u,0)$ is completed. Event A_2 comes under coincidence of the following independent events: **1)** at first takes place change of phase on failures (the failure comes really, but QS don't be transferred to recovery ($\alpha\, exp(-\alpha u)\, du$); **2)** for a time $x+u$ transmission of j-th block should not be completed $F(x+u)$; **3)** for a time $x+u+\nu$ transmission of j-th block will be completed under condition, that it should be completed for a time $x+u$: $d_\nu F(x+u+\nu)/\bar{F}(x+u)$; **4)** for a time ν QS will pass $pl+k+1$ phases on failures ($B_*^{(pl+k-1)}(\upsilon)-B_*^{(pl+k)}(\upsilon)$, $\quad P=\overline{0,\infty}$, $\quad x=\overline{1,l}$), **5)** for a time $t-u-\nu$ transmission of the remained part of message will be completed starting from j-th block (j-th block is transmissed repeatedly) in k state (at the beginning of message transmission) on failures and quantity of transmissions equals to $\nu+1$ $\left(i=\overline{1,r-1}\right)$

Denote

$$\overline{\Psi}_j^{(kv)}(s,x)=\int_0^\infty e^{-st}\psi_j^{(kv)}(t,x)dt;\quad \overline{\Phi}_j^{(kv)}(s,x)=\int_0^\infty e^{-st}\Phi_j^{(kv)}(t,x)dx;$$

$$\varphi_1(s)=\int_0^\infty e^{-su}d_uF(x+u)=e^{-s(\tau_b-x)};$$

$$\varphi_2(s)=\int_0^\infty e^{-sy}dy\int_0^y\left[B_*^{(pl+k-1)}(y-u)-B_*^{(pl+k)}(y-u)e^{-\alpha u}du=\right.$$

$$=\frac{1}{(s+\alpha)^2}\left(\frac{\alpha}{s+\alpha}\right)^{pl+k-1}$$

and taking the Laplace-Stilties transform to (1), (2), (3), we'll obtain

$$\frac{d\overline{\Psi}_j^{(kv)}(s,x)}{dx}-(s+\alpha)\overline{\Psi}_j^{(k,v)}(s,x)+\alpha\overline{\Psi}_j^{(k+1,v)}(s,x)=0, \tag{4}$$

$$\overline{\Psi}_j^{(kv)}(s,0)=\overline{\Phi}_j^{(kv)}(s,0),\quad j=\overline{1,n},\ k=\overline{1,l-1},\ v=\overline{1,r};$$

$$\overline{\Psi}_j^{(lv)}(s,x)=e^{-(\alpha+s)(\tau_b-x)}\overline{\Phi}_{j+1}^{(l1)}(s,0)+$$

$$+\sum_{k=1}^{l}\sum_{p=0}^{\infty}\alpha\overline{\Phi}_j^{(k,v+1)}(s,0)\frac{1}{2\pi i}\int_{C^*-i\infty}^{C^*+i\infty}\varphi_1(s-w)\varphi_2(w)dw; \tag{5}$$

$$\overline{\Psi}_j^{(l,r)}(s,x)=e^{-(\alpha+s)(\tau_b-x)}\overline{\Phi}_{j+1}^{(l1)}(s,0)+g(s)\overline{\Phi}_j^{(1,1)}(s,0)\left(e^{-s(\tau_b-x)}-e^{-(\alpha+s)(\tau_b-x)}\right); \tag{6}$$

$$j=\overline{1,n}$$

The Laplace transform for the boundary conditions has the form $\overline{\Phi}_{n+1}^{(l1)}(s,0)=\frac{1}{s}$.

Here C^* - an abscissa of convergence of improper integral which lies into field of analyticity of subintegral function: $i=\sqrt{-1}$. Substituting of $\varphi_1(s)$ and $\varphi_2(s)$ in the expression (5) and after its calculating with the help of deductions, we'll obtain:

$$\overline{\Psi}_j^{(lv)}(s,x)=e^{-(\alpha+s)(\tau_b-x)}\overline{\Phi}_{j+1}^{(l1)}(s,0)+$$

$$+\sum_{k=1}^{l-1}\left[\overline{\Phi}_j^{(k,v+1)}(s,0)\alpha^k\sum_{p=1}^{l}e^{-(s-w_p)(\tau_b-x)}(w_p+\alpha)^{l-1-k}\Big/\left(\prod_{\eta=l,\eta\neq p}^{l}(w_p-w_\eta)\right)\right]+$$

$$+\left[\overline{\Phi}_j^{(l,v+1)}(s,0)\alpha^l\sum_{p=0}^{l}e^{-(s-w_p)(\tau_b-x)}\Big/\left(\prod_{\eta=0,\eta\neq p}^{l}(w_p-w_\eta)\right)\right] \tag{7}$$

Where $w_d=\alpha(e^{2\pi di/l}-1)$, $d=\overline{1,l}$ - zeroes of equations $(w+\alpha)^l-\alpha^l=0$, $w_0=-\alpha$.

With the purpose of further calculations we introduce a new variable $y=\tau_b-x$ (where y - the rest time before finishing of block transmission). Substituting in them $y=\tau_b-x$ into the equations (4), (5) and (7) and taking the Laplace transform on argument y (corresponding operator ω), we obtain:

$$(\omega+s+\alpha)\tilde{\overline{\psi}}_j^{(kv)}(s,\omega)=\overline{\Phi}_{j+1}^{(kv)}(s,0)+\alpha\tilde{\overline{\Psi}}_j^{(k+1,v)}(s,\omega), \tag{8}$$

$$j=\overline{1,n},\ k=\overline{1,l-1},\ n=\overline{1,r};$$

$$\tilde{\Psi}_j^{(l\nu)}(s,\omega)=\Phi_{j+1}^{(ll)}(s,0)/(\omega+s+a)+$$

$$+\sum_{k=1}^{l-1}\left\{a^k\Phi_j^{(k,\nu+1)}(s,0)\sum_{p=1}^{l}\left[\left(w_p+a\right)^{l-1+k}\Big/\prod_{\eta=1,\eta\neq p}^{l}\left(w_p-w_\eta\right)\right]\Big/\left(\omega+s-w_p\right)\right\}+$$

$$+a^l\Phi_j^{(l,\nu+1)}(s,0)\sum_{p=0}^{l}\left\{\left[1\Big/\prod_{\eta=1,\eta\neq p}^{l}\left(w_p-w_\eta\right)\right]\Big/\left(\omega+s-w_p\right)\right\},\quad (9)$$

$$\nu=\overline{1,r-1},\ w_d=a\left(e^{i2pd/l}-1\right),\ d=\overline{1,l},\ w_0=-\alpha;$$

$$\tilde{\Psi}_j^{(lr)}(s,\omega)=\Phi_{j+1}^{(ll)}(s,0)/(\omega+s+\alpha)+\{a\bar{g}(s)/[(s+\omega)(s+\omega+\alpha)]\}\,\Phi_j^{(l,l)}(s,0);\quad (10)$$

$$\Phi_{n+1}^{(k\nu)}(s,0)=\frac{1}{s}\quad (11)$$

Here

$$\Psi_j^{(k\nu)}(t,x)=\Psi_j^{(k\nu)}(t,\tau_b-y)=\tilde{\Psi}_j^{(k\nu)}(t,y);\ \tilde{\bar{\Psi}}_j^{(k\nu)}(s,y)=\int_0^\infty e^{-st}\tilde{\Psi}_j^{(k\nu)}(t,y)dt;$$

$$\tilde{\Psi}_j^{(k\nu)}(t,0)=\Psi_j^{(k\nu)}(t,\tau_b)=\Phi_{j+1}^{(k\nu)}(t,0);\ \tilde{\bar{\Psi}}_j^{(k\nu)}(s,0)=\bar{\Phi}_{j+1}^{(k\nu)}(s,0);$$

$$\tilde{\bar{\Psi}}_j^{(k\nu)}(s,\omega)=\int_0^\infty e^{-\omega y}\tilde{\bar{\Psi}}_j^{(k\nu)}(s,y)dy$$

in consequence of solution of system of algebraic equations (8), (9) and (10) system, with account (11), we find $\tilde{\bar{\Psi}}_j^{(k\nu)}(s,\omega)$ on operator ω, we define $\tilde{\bar{\Psi}}_j^{(k\nu)}(s,y)$, and after substituting $y=0$ in it ($x=\tau_b$), we define meaning $\tilde{\bar{\Phi}}_{j+1}^{(k\nu)}(s,0)$, ($k=\overline{1,l}$, $n=\overline{1,r}$, $j=\overline{1,n}$).

For example: let l=2, r=2, n=1 (j=1), $\Phi_{z2}^{(k\nu)}(s,0)=1/s$

In conformity with (8):

$$\tilde{\bar{\Psi}}_j^{(1,i)}(s,\omega)=\frac{1+s\alpha\tilde{\bar{\Psi}}_1^{(2,i)}(s,\omega)}{s(s+\omega+a)},\quad i=1,2$$

In conformity with (10):

$$\tilde{\bar{\Psi}}_1^{(2,2)}(s,\omega)=\frac{1}{s(s+\omega+\alpha)}+\left[\frac{\alpha g(s)}{(s+\omega)(s+\omega+\alpha)}\right]\bar{\Phi}_1^{(1,1)}(s,0)$$

In conformity with (9):

$$\tilde{\bar{\Psi}}_1^{(2,1)}(s,\omega)=\frac{1}{s(s+\omega+\alpha)}+\alpha\bar{\Phi}_1^{(1,2)}(s,0)\times$$

$$\times\left[\frac{1}{(w_1-w_2)(s+\omega+w_1)}+\frac{1}{(w_2-w_1)(s+\omega-w_2)}\right]+\alpha^2\bar{\Phi}_1^{(2,2)}(s,0)\times$$

$$\times\left[\frac{1}{(w_0-w_1)(w_0-w_2)(s+\omega+w_0)}+\frac{1}{(w_1-w_0)(w_1-w_2)(s+\omega+w_1)}+\right.$$

$$\left.+\frac{1}{(w_2-w_0)(w_2-w_1)(s+\omega-w_2)}\right];$$

Considering:

$$w_0=-\alpha,\ w_1=\alpha\left(e^{i2\pi/2}-1\right)=-2\alpha,\ w_2=\alpha\left(e^{2\pi i}-1\right)=0$$

Latter expression can be presented in the following form

$$\tilde{\tilde{\Psi}}_1^{(2,1)}(s,\omega)=\frac{1}{s(\omega+s+\alpha)}+\frac{\alpha\overline{\Phi}_1^{(1,2)}(s,0)}{(\omega+s)(\omega+s+2\alpha)}+\frac{\alpha^2\overline{\Phi}_1^{(2,2)}(s,0)}{(\omega+s+\alpha)(\omega+s+2\alpha)(\omega+s)} \qquad (12)$$

After going from ω to y in (12) we obtain:

$$\tilde{\overline{\Psi}}_1^{(2,1)}(s,y)=\frac{1}{s}e^{-(s+\alpha)y}+\frac{e^{-sy}-e^{-(s+2\alpha)y}}{2}\overline{\Phi}_1^{(1,2)}(s,0)+$$

$$+\left[\frac{1}{2}e^{-sy}+\frac{1}{2}e^{-(s+2\alpha)y}-e^{-(s+\alpha)y}\right]\overline{\Phi}_1^{(2,2)}(s,0) \qquad (13)$$

Supposing $y=\tau_b$ in (13) and considering $\tilde{\overline{\Psi}}_1^{(2,1)}(s,\tau_b)=\overline{\Phi}_1^{(2,1)}(s,0)$, we obtain:

$$\overline{\Phi}_1^{(2,1)}(s,0)=\frac{1}{s}e^{-(s+\alpha)\tau_b}+\frac{1}{2}\left(e^{-s\tau_b}-e^{-(s+2\alpha)\tau_b}\right)\overline{\Phi}_1^{(1,2)}(s,0)+$$

$$+\left[\frac{1}{2}e^{-s\tau_b}+\frac{1}{2}e^{-(s+2\alpha)\tau_b}-e^{-(s+\alpha)\tau_b}\right]\overline{\Phi}_1^{(2,2)}(s,0)$$

Introducing the corresponding denotation we obtain:

$$\Phi_1^{(2,1)}(s,0)=A(s,T)+B(s,T)\,\overline{\Phi}_1^{(1,2)}(s,0)+C(s,T)\,\overline{\Phi}_1^{(2,2)}(s,0);$$

Analogously:

$$\overline{\Phi}_1^{(2,2)}(s,0)=A(s,T)+D(s,T)\,\overline{\Phi}_1^{(1,1)}(s,0);$$

$$\overline{\Phi}_1^{(1,1)}(s,0)=A(s,T)+E(s,T)\,\overline{\Phi}_1^{(2,1)}(s,0);$$

$$\overline{\Phi}_1^{(1,2)}(s,0)=A(s,T)+F(s,T)\,\overline{\Phi}_1^{(2,2)}(s,0);$$

Solving four equations with four unknown values, we define:

$$\overline{\Phi}_1^{(1,1)}(s,0),\ \overline{\Phi}_1^{(1,2)}(s,0),\ \overline{\Phi}_1^{(2,1)}(s,0),\ \overline{\Phi}_1^{(2,2)}(s,0)$$

Then we find:

$$\overline{\Phi}(s,T)=\overline{\Phi}_1^{(1,1)}(s,0)+\overline{\Phi}_1^{(1,2)}(s,0)+\overline{\Phi}_1^{(2,1)}(s,0)+\overline{\Phi}_1^{(2,2)}(s,0)$$

and $(s)=\int_0^\infty d_u F(u)\Phi(s,u)$, here $\Phi(s,u)$ is obtained from $\Phi(s,T)$ by replacement T on u.

REFERENCES

1. *V.Shvartsman, G.Emelianov.* Teoria Peredachi Diskretnoi Informatsii. M. 1979 (Russian).
2. *I.Mizin, A.Urinson, G.Khromeshin.* Peredacha Informatsii v Setiakh s Kommutatsiei Soobshenii. M. 1977 (Russian).
3. *L.Kleinrok.* Teoria Massovogo Obslujivania. M. 1979 (Russian).
4. *I.Mikadze.* J. Avtomatika i Telemekhanika. **2**, 1979. 173-186.
5. *I.Mikadze.* J. Kibernetika. **1**, 1988. 56 - 61.
6. *R.Kakubava, I.Mikadze et al.* J. Avtomatika i Telemekhanika. **7**, 1988. 176 - 184.
7. *R.Kakubava, I.Mikadze.* J. Avtomatika i Telemekhanika. **1**, 1988.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

AUTOMATIC CONTROL AND
COMPUTER ENGINEERING

L. Mirtskhulava

On the Data-Transmission Channel Models with Account Information Distortion

Presented by Academician V.Chavchanidze, September 9, 1996

ABSTRACT. **On the basis of the queueing theory the problem of determination of the probability characteristic of time of message transmission in data-transmission channel with account information distortion is considered in the paper.**

On the different stages of the design, installation and modernization of information nets (IN) it is necessary to investigate information processing proceeding in the nets with the purpose of appraisal of its operation quality indices and the basis of adopted engineering decisions. IN complication doesn't often allow to build models which are completely adequate to real processes. It is necessary to determine degree of model adequacy of IN and consequently the degree of trust to the IN investigation results with the help of the model.

If investigated model of IN has functionally given and if IN may be broken into parts so that the information exchange at their joint will take place without space and time distortions; then each component can be presented as a unit. Modelling of IN can be reduced to modelling of each unit separately with following unification of these models into the unit system where the outputs of one model are switched on to the entrances of other unit on given beforehand rules. Under modelling of IN it is necessary to have the following totality of models: data - transmission channel (DTC) model; data - transmission path model; switching centre and information net models.

The creation of DTC model, which reflects fully enough the processes proceeding in real channels is very important stage of the work in common complex of IN modelling tasks so far as quality of DTC model determines beforehand the degree of adequacy of model creation to the processes following in the real net.

Three mathematical models of data - transmission channel assuming that in DTC failures can occur essentially influencing on reliability and data - transmission truthworthness are considered in the paper.

An analytical expression of time distribution function of the message transmissions which consists of arbitrary number of information blocks has been received for each of the models.

The following denotations are introduced in the paper:

n - number of information blocks in message.

$\Phi_j(t)$ - the distribution function of the message transsmission, time probability which consists of n information blocks if transmission is started from j-th block ($j=\overline{1,n}$).

$F_j(u)$ - the distribution function of transmissed block length.

$r-1$ - the quantity of information block transmission repeatition on DTC before repair.

α – the traffic rate of errors origin.

$G(v)$ – the distribution function of recovery.

Under creation of these models given in this article we used the following assumptions:

the channel is considered unreliable; the time is distributed exponentially between the neighbouring errors;

the transmissed message on DTC consists of n information blocks with constant length distributed by law $F_j(u)=1(t-\tau_j)$, $j=\overline{1,n}$, where τ_j– j-th block transmission time;

the check of accuracy of getting information is made immediately after receiving the next block and the time spent on check is neglectly small, check is full.

MODEL 1. Under assumptions (A) in case of error origin (information distortion) block transmission is repeated until receiving accurate information but no more than r–1 times. The channel is transferred on check after r-times repetition. Then the information transmission is renewed from the first block transfer.

The DTC work on information transmission, starting from j-th block ($j=\overline{1,n}$), is discribed by following integral equations:

$$\Phi_j^{(i)}(t)=\int_0^t dF_j(u)e^{-\alpha_j u}\Phi_{j+1}^{(1)}(t-u)+\int_0^t dF_j(u)\left[1-e^{-\alpha_j u}\right]\Phi_j^{(i+1)}(t-u) \qquad (1)$$

$$\Phi_j^{(r-1)}(t)=\int_0^t dF_j(u)e^{-\alpha_j u}\Phi_{j+1}^{(1)}(t-u)+\int_0^t dF_j(u)\left(1-e^{-\alpha_j u}\right)\int_0^{t-u} dG(\upsilon)\Phi_1^1(t-u-\upsilon) \qquad (2)$$

$$j=\overline{1,n} \qquad \Phi_{n+1}(t)=1 \qquad (3)$$

Using the Laplace transform, we obtain:

$$\overline{\Phi}_j^i(s)=\overline{f_j}\left(s+\alpha_j\right)\overline{\Phi}_{j+1}^{(1)}(s)+\left[\overline{f_j}(s)-\overline{f_j}\left(s+\alpha_j\right)\right]\overline{\Phi}_j^{(i+1)}(s); \qquad (4)$$

$$\overline{\Phi}_j^{(r-1)}(s)=\overline{f_j}\left(s+\alpha_j\right)\overline{\Phi}_{j+1}^{(1)}(s)+\left[\overline{f_j}(s)-\overline{f_j}\left(s+\alpha_j\right)\right]\overline{g}(s)\overline{\Phi}_1^{(1)}(s); \qquad (5)$$

$$\overline{\Phi}_{n+1}(s)=\frac{1}{s}; \qquad \overline{f_j}(s)=\int_0^\infty e^{-st}dF_j(t); \qquad (6)$$

$$\overline{\Phi}_j^{(i)}(s)=\int_0^\infty e^{-st}\Phi_j^{(i)}(t)dt; \qquad (7)$$

$$\overline{g}(s)=\int_0^\infty e^{-st}dG(t); \qquad j=\overline{1,n}; \qquad i=\overline{1,r-2} \qquad (8)$$

The solutions of these equations have the form:

$$\overline{\Phi}_1^{(1)}(s)=\frac{a^n(s)[a(s)-1]}{s[(a(s)-1)-(a^n(s)-1)b(s)]} \tag{9}$$

The mean time of task execution T_m is calculated by the following equation:

$$-T_m=s\overline{\Phi}_1^{(1)}(s)\Big|_{s=0}^{1}=\frac{a^n(s)[a(s)-1]}{[(a(s)-1)-(a^n(s)-1)b(s)]}\Bigg|_{s=0}^{1} \tag{10}$$

where

$a(0)=1-b(0);\quad b(0)=[1-\bar{f}(\alpha)]^{r-1};$

$\tau_r=-g'(0);\quad \tau_b=-f'(0)$

$$-T_m=\frac{\tau_b+[1-\bar{f}(\alpha)]^{r-1}[\tau_r\bar{f}(\alpha)-\tau_b]}{\bar{f}(\alpha)a^n(0)b(0)}-\frac{\tau_b+[1-\bar{f}(\alpha)]^{r-1}[\tau_r\bar{f}(\alpha)-\tau_b]}{\bar{f}(\alpha)b(0)};\ r\geq 3 \tag{11}$$

MODEL 2. Under assumptions (A) in the case of information distortion block transmission is repeated until it won't be correctly received. It takes place with a help of control categories.

Analogously to previous model the given model is described by the following system of equations:

$$\Phi_j^{(i)}(t)=\int_0^t dF_j(u)e^{-\alpha u}\Phi_{j+1}^{(1)}(u)+\int_0^t dF_j(u)\left[1-e^{-\alpha u}\right]\int_0^{t-u} dG(\upsilon)\Phi_j^{(1)}(t-u-\upsilon) \tag{12}$$

$j=\overline{1,n};\quad \Phi_{n+1}(t)=1$

The solution has the form:

$$\overline{\Phi}_1^{(1)}(s)=\frac{\bar{f}^n(s+\alpha)}{\{1-[\bar{f}(s)-\bar{f}(s+\alpha)]g(s)\}^n}\frac{1}{s} \tag{13}$$

$$-T_m=n\frac{\tau_b+\tau_r[1-\bar{f}(\alpha)]}{\bar{f}(\alpha)} \tag{14}$$

MODEL 3. Under assumptions (A) in the case of the arrising errors (information distortion) the block transmission is repeated until the receiving of accurate information but no more than $r-1$ times. The channel is transferred for repair after r times repetition. The information transmission is renewed after the repair from distorted block repetition.

The DTC work on information transmission, starting from j-th block ($j=\overline{1,n}$), is described by the following integral equations:

$$\Phi_j^{(i)}(t) = \int_0^{t} dF_j(u) e^{-\alpha_j u} \Phi_{j+1}^{(1)}(t-u) + \int_0^{t} dF_j(u)\left[1 - e^{-\alpha_j u}\right] \Phi_j^{(i+1)}(t-u), \quad (15)$$

$$\Phi_j^{(r-1)}(t) = \int_0^{t} dF_j(u) e^{-\alpha_j u} \Phi_{j+1}^{(1)}(t-u) + \int_0^{t} dF_j(u)\left[1 - e^{-\alpha_j u}\right] \int_0^{t-u} dG(\upsilon) \Phi_j^{(1)}(t-u-\upsilon), \quad (16)$$

$j = \overline{1,n}$; $\Phi_{n+1}(t) = 1$

The solution has the form:

$$\Phi_1^{(1)}(s) = \frac{a^n(s)}{s[1-b(s)]^n} \quad (17)$$

$$-T_m = s\Phi_1^{(1)}(s)\Big|_{s=0}^{1} \quad (18)$$

For example:

Let assume that the message is transmissed with constant length $-T_m$, then

$$\tau_b = \frac{T_m}{n} + \tau_c \ ; \quad \bar{f}(\alpha) = e^{-\tau_b \alpha}$$

where n - quantity of blocks; τ_c - time of check categories transfer, accompaning each block plus time of check in digit computer (DC) (the latter may be neglected); i - quantity of repetitions. Under $i \le r$ (r = 2, 3,...) the channel is transferred for repair; τ_r - mean time of channel recovery, it is obvious, that

$$\tau_c = \tau_b \frac{\tilde{n}_{c_c}}{m_{bc}} + \tau_d$$

$$\alpha = \frac{1}{kc_{ch}}$$

$\tilde{n}$ - quantity of categories in course of transfer, in which one error arrises averagely, and $\bar{\bar{c}}_c$ - speed of transfer (channel productivity)

$$T_m = N_m \cdot c_c$$

N_m - quantity of categories (bits) in message;

c_c - transfer time of one bit.

REFERENCES

1. *V. O. Shvartsman, G. A. Emelianov.* Teorija Peredachi Diskretnoj Informacii. M., 1979.

2. *I.A.Mizin, L.S.Urinson, G.K.Khrameshin.* Peredacha Informacii v Setijakh s Kommutatciej Soobshenii. 1, 1977.
3. *L.Kleinrok.* Teorija Massovogo Obslujivania. 1, 1979.
4. *I.Mikadze.* Avtomatika i Telemekhanika, 2, 1979, 173-186 (Russian).
5. *Idem.* Kibernetika. 1.1988, 56-61 (Russian).
6. *R.Kakubava, I.Mikadze et al.* Avtomatika i Telemekhanika, 7, 1988, 176-184.
7. *R.Kakubava, I.Mikadze.* Avtomatika i Telemekhanika 1, 1988 (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

AUTOMATIC CONTROL AND
COMPUTER ENGINEERING

T.Pestvenidze, M.Karkashadze

Fulfilling the Task in the Redundant Engineering Systems

Presented by Academician V.Chichinadze, September 2, 1996.

ABSTRACT. **The work considers probabilities of fulfilling the task of the redundant complex joint usage engineering system with a multistate working ability within the given time.**

One of the main aspects of the effective functioning of the engineering systems involving unreliable equipments in their structure is the minimization of the unproductive losses of time, connected with the expense of time or restoration and switching into the reserve apparatus after the failures observing the failures, work repetition devaluated by the failures, expectation of loading in the workable state. In many systems the constant control is attainable over the working ability of the equipments allowing to observe any failure at the moment of their origin.

Besides, very frequently the failures can be considered non-devaluating. It means that the system after the restoring was interrupted. That's why all the production between the neighbouring failures appears useful. In the present work we consider homogeneous multiprocess (multimachine) system (MPCS) as the object of our research.

Let MPCS consisting of $\overline{m}$ main and $n-m$ reserve equipments serve the demand (task) of the constant volume E (with the distributing function $F(t) = 1(t - \tau_3)$. Here τ_3 is the time of fulfilling the task of the ideal system i.e. the system without failures. The task is distributed among all the processing MPCS equipments. (It's supposed that the structure and the algorithms of the MPCS make possible to break the task into the parallel branches). All the MPCS equipments are completely intersubstitutable and they process the demand in the mode of mutual assistance. At unfailing work the task can be fulfilled by one equipment within the time $T_3 = \dfrac{E}{C}$; and by the equipment i– within the time $u = \dfrac{E}{cf(i)} = \dfrac{\tau_3}{f(i)}$ (Here C–the nominal productivity of each MPCS equipment. The kind of the function $f(i)$ $(1 \leq f(i) \leq i)$ is determined by the expenditure of the resources (apparatus and time) on complexing while fulfilling i $(i = \overline{1,m})$ by the workable equipments of the general task volume E [1,2]. The working equipments fail with the intensity B_1 and the reserve equipments – with the intensity B_2. The restoration time is an occasional quantity with the demonstrative distribution with the parameter M_1. The failures of the separate equipments don't devaluate already fulfilled work, if even only one of the equipments is workable (thus the given model differs from the analogical model considered in [12]), if by the moment of the failure all the equipments had been in the failing state then after the restoring the recounting of the distorted parts takes place (the fulfilled work is partialy devaluated in the limits of the block – stage j). The function of the distribution MPCS reconfiguration time (turning the system from the state i into the state j) is the occasional quantity with the distributing function $G_{ij}(v)$. The other terms of functioning of the given model in the busy mode completely coincide with its functioning in the free mode.

We designate by $H_{ij}^{(l)}(t, x)$ – the probability of completing the processing of the package which consists of the l demands within the time less than t MPCS– being in the state j on the term that the fulfilling of the task was renewed at the moment of the time
$t = 0$, when the system was in the state i and for completing the serving of the demands package the required time was $y = \tau_3 - x$,. at the unfailing work of one of its equipments (i.e. recounting on one equipment).

The aim of our work is to determine the Laplace transformation of the Function $H_{ij}^{(l)}(t, x)$. Further we'll designate it by $H_{ij}(t, x)$. It is determined through the following equation system.

$$\overline{F}(x)H_{ij}(t,x) = \delta_{ij}\int_0^t \exp(-c_i^0 u)\, d_uF(x+n_i u) +$$

$$+ (1-\delta_{in})\mu_l \int_0^t \exp(-c_i^0 u)\, \overline{F}(x+n_i u)du \int_0^{t-u} H_{i+1,j}(t-u-v, x+n_i u)\, dG(v) +$$

$$+ c_i \int_0^t \exp(-c_i^0 u)\, \overline{F}(x+n_j u)du \int_0^{t-u} H_{i-1,1}(t-u-v, x+n_j u)\, dG_{i,i-1}(v), \qquad (1)$$

$$i = \overline{2,n}\,,\; j = \overline{1,n}$$

$$\overline{F}(x)H_l(t,x) = \delta_{ij}\int_0^t \exp(-c_l^0 u)du F(x+u) +$$

$$+ (1-\delta_{ln})\mu_l \int_0^t \exp(-c_l^0 u)\, \overline{F}(x+u)du \int_0^{t-u} H_{2j}(t-u-v, x+u)\, dG_{l2}(v) +$$

$$+ \beta_l \int_0^t \exp(-c_l^0 u)\, \overline{F}(x+u)du \int_0^{t-u} \mu_l \exp(-\mu_l v)\, dv \int_0^{t-u-v} H_{lj}(t-u-v-v, 0)dG_{0l}(v), \qquad (2)$$

$$j = \overline{1,n}$$

Here

$$c_i = i\beta_l\delta_0(i<m) + [m\beta_l + (i-m)\beta_2]\delta_l(i \ge m);$$
$$c_i^0 = (1-\delta_{in})\mu_l + c_j;\; n_i = \delta_0(i<m)f(i) + \delta_l(i \ge m)f(m);$$
$$\overline{F}(x) = 1 - F(x).$$
$$\delta(\cdot) = \begin{cases} 1, \text{if}(\cdot) \text{ is a true statement} \\ 0, \text{ on the contrary} \end{cases}$$

δ_{ij}, δ_{in} – Kroneker symbols.

The initial and the boundary terms have the form: $H_{ij}(0, x) = 0$, when

$$x \neq \tau_3 (i,j = \overline{1,n})$$

$$H_{ij}(0, x) = \begin{cases} \delta_{ij}, \text{when } x = \tau_3, \\ 0, \text{when } x > \tau_3; \end{cases}$$

$$i, j = \overline{1,n}\,;\; t = \overline{0,\infty}.$$

As an example we clarify the second member of equation (1), when $i > m$ $(i \neq n)$. It's the common probability of the following:

1) The restoration of one of the $n - i$ equipments in disrepair being at the $t = 0$ moment in restoration will be completed within the time interval $(u, u + du)$.

2) Within the time u the serving of the processed package of demands won't be completed and neither the workers nor the reserve equipments will fail.

3) For the reconfiguration of the serving system (SS) is required time V.

4) The serving of the package will be completed CS being in the state j within the time less than $t - u - v$ if at the moment $t = u + v$ the system is in the state $i + 1$ and for the completing the processing of the demands package $\tau_3 - x - mu$ time is required at the constant and unfailing work of one equipment.

Designating

$$\Phi_{ij}(t, x) = \overline{F}(x) H_{ij}(t, x),\; \overline{\Phi}_{ij}(s, x) = \int e^{-st} \Phi_{ij}(t, x)dt,$$

$$\overline{g}_{ij}(s) = \int_0^\infty e^{-st} \overline{g}_{ij}(t)dt,\; \overline{g}_{ij}(t) = G_{ij}'(t),\; \overline{f}(s) = \int_0^\infty e^{-st} dF(u) = e^{-s\tau_3},$$

$$\overline{H}_{ij}(s) = \int_0^\infty e^{-st} H_{ij}(t)dt$$

and using the Laplace transformation (1) and (2) after a simple transformation we get:

$$\exp\left[-\left(s + c_i^0\right)x / n_i\right]\overline{\Phi}_{ij}(s, x) = \delta_{ij} \exp\left[-\left(s + c_i^0\right)\tau_3 / n_i\right] / s +$$

$$+ [(1 - \delta_{in})\mu_1 \overline{g}_{i,i+1}(s)]/n_i \int_x^{\tau} \exp\left[-\left(s + c_i^0\right)\tau / n_i\right]\overline{\Phi}_{i+1,i}(s, \tau)d\tau + \tag{3}$$

$$+ [c_i \overline{g}_{i,i-1}(s)/n_i] \int_x^{\tau} \exp\left[-\left(s + c_i^0\right)\tau / n_i\right]\overline{\Phi}_{i-1,j}(s, \tau)d\tau,\; i = \overline{2,n},\; j = \overline{1,n};$$

$$\exp\left[-\left(s + c_i^0\right)x\right]\overline{\Phi}_{1j}(s, x) = \delta_{1j} \exp\left[-\left(s + c_i^0\right)\tau_3\right] / s +$$

$$+ [(1 - \delta_{1n})\mu_1 \overline{g}_{12}(s) \int_x^{\tau_\delta} \exp\left[-\left(s + c_i^0\right)\tau\right]\overline{\Phi}_{2j}(s, \tau)d\tau + \tag{4}$$

$$+ \beta_1 \overline{g}_{01}(s)\mu_1 \overline{H}_{1j}(s, 0)\left[\exp\left(-c_1^0 - s\right)x - \exp\left(-s - c_1^0\right)\tau_\delta\right] / (s + \mu_1)\left(s + c_1^0\right).$$

Through the differentiation of both parts (3) and (4) according to the argument x we get:

$$\overline{\Phi}_{ij}(s, x) - \left[\left(s + c_i^0\right) / n_i\right]\overline{\Phi}_{ij}(s, x) + \left[(1 - \delta_{in})\mu_1 \overline{g}_{i,i+1}(s) / n_i\right]\overline{\Phi}_{i+1,j}(s, x) + \tag{5}$$

$$+\left[c_i\bar{g}_{i,i-1}(s)/n_i\right]\bar{\Phi}_{i-1,j}(s,x)=0,\ i=\overline{2,n};$$

$$-\left(c_1^0+s\right)\bar{\Phi}_{1j}(s,x)+\bar{\Phi}_{1j}'(s,x)=(1-\delta_{1n})\mu_1\bar{g}_{12}(s)\bar{\Phi}_{2j}(s,x)- \tag{6}$$

$$-\beta_1\bar{g}_{01}(s)\mu_1\bar{H}_{1j}(s,0)/(s+\mu_1),\ j=\overline{1,n}$$

We designate by y the time required to complete the serving of the demands begun at the moment $t=0$. Substituting τ_3-y in (3) and (4) with x and turning to the laplace transformation according to the argument y (with account $\bar{\Phi}_{ij}^0(s,y)=\delta_{ij}/s$, when $y=0$ and $\bar{\Phi}_{ij}^0(s,y)=0$, when $y<0$) we get:

$$-\left[w\bar{\bar{\Phi}}_{ij}^0(s,w)-\delta_{ij}/s\right]-\left[\left(s+c_i^0\right)/n_i\right]\bar{\bar{\Phi}}_{ij}^0(s,w)+\left[\left(1-\delta_{in}\right)\mu_1\bar{g}_{i,i+1}(s)/n_j\right]\bar{\bar{\Phi}}_{i+1,j}^0(s,w)+$$

$$+\left(c_i\bar{g}_{i,i-1}(s)/n_i\right)\bar{\bar{\Phi}}_{i-1,j}^0(s,w)=0,\ i=\overline{2,n}; \tag{7}$$

$$-(w+s+c_1^0)\bar{\bar{\Phi}}_{1j}^0(s,w)+(1-\delta_{1n})\mu_1\bar{g}_{12}(s)\bar{\bar{\Phi}}_{2j}^0(s,w)=$$

$$=-\delta_{1j}/s-[\beta_1\mu_1\bar{g}_{01}(s)/w(s+\mu_1)]\bar{H}_{1j}(s,0);\ j=\overline{1,n} \tag{8}$$

Here

$$\bar{\Phi}_{ij}^0(s,y)=\bar{\Phi}_{ij}^0(s,\tau_3-y);\ \bar{\bar{\Phi}}_{ij}^0(s,w)=\int_0^\infty e^{-wy}\,\bar{\Phi}_{ij}^0(s,y)\,dy.$$

Using the Kronneker symbol we can demonstrate (7) and (8) in the following form:

$$a_i(s)\bar{\bar{\Phi}}_{i-1,j}^0(s,w)+b_i(s,w)\bar{\bar{\Phi}}_{ij}^0(s,w)+d_i(s)\bar{\bar{\Phi}}_{i+1,j}^0(s,w)=F_{ij}(s); \tag{9}$$

$$i,j=\overline{1,n};\ \bar{\bar{\Phi}}_{0j}^0(s,w)=0.$$

Here

$$a_i(s)=c_i\bar{g}_{i,i-1}(s)/n_i;\ b_i(s,w)=-\left[w+\left(s+c_i^0\right)/n_i\right];$$

$$d_i(s)=(1-\delta_{in})\mu_1\bar{g}_{i,i+1}(s)/n_i;\ F_{ij}(s)=-\delta_{ij}/s-[\delta_{1i}\beta_1\mu_1\bar{g}_{01}(s)/w(s+\mu_1)]\bar{H}_{1j}(s,0).$$

Solution (9) can be obtained through the determinants D_n and Δ_n of the submatrices of the matrix $D(s,w)$ [3].

$$D(s,w)=\begin{bmatrix} b_1 & d_1 & 0 & 0 & \cdots & 0 & 0\\ a_2 & b_2 & d_2 & 0 & \cdots & 0 & 0\\ 0 & a_3 & b_3 & d_3 & \cdots & 0 & 0\\ \cdots & \cdots & \cdots & \cdots & \cdots & \cdots & \cdots\\ 0 & 0 & 0 & 0 & \cdots & a_n & b_n \end{bmatrix};$$

$$D_0(s,w)=1;\ D_1(s,w)=b_1;\ D_m=b_mD_{m-1}-a_md_{m-1}D_{m-2},\ m=\overline{2,n};$$

$$\Delta_0=1;\ \Delta_1=b_n;\ \Delta_m=b_{n-m+1}\Delta_{m-1}-a_{n-m+2}d_{n-m+1}\Delta_{m-2}$$

Here the arguments of the coefficients are omitted.

The solution has the form:

$$\overset{=0}{\Phi}_{ij}(s,w)=\left[\Delta_{n-i}(s,w)\sum_{\sigma=1}^{i}(-1)^{\sigma+1}F_{\sigma j}(s)D_{\sigma-1}(s,w)\prod_{\eta=\sigma}^{i-1}a_{\eta+1}+\right.$$

$$\left.+D_{i-1}\sum_{\sigma=i+1}^{n}F_{\sigma j}(s)\Delta_{n-\sigma}(s,w)\prod_{\eta=i}^{\sigma-1}(-d_{\eta})\right]/D_n;\ i,j=\overline{1,n}.$$

After the reserve transformation $\overset{=0}{\Phi}_{ij}(s,w)$ according to the operator w we find $\overset{=0}{\Phi}_{ij}(s,y)$ and accordingly $\overline{\Phi}_{ij}(s,x)$.

It's evident that $\overline{H}_{ij}(s)=\overline{\Phi}_{ij}(s,0)$ and $\overline{H}_i(s)=\sum_{j=1}^{n}\overline{\Phi}_{ij}(s,0)$.

Georgian Technical University

REFERENCES

1. *I.S.Mikadze.* J. Kibernetika, 3, 1989.
2. *R.V.Kakubava.* Thesis for Doctor's Degree. Tbilisi, 1994.
3. *Drh.Riordan.* Verojatnostnye Sistemy Obsluzhivanya. M., 1996, (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

SOIL SCIENCE

V.Lomidze

Peculiarities of Temperature Regime in Brown Soils

Presented by Corr. Member of the Academy T.Urushadze, August 15, 1996

ABSTRACT. **This report presents the temperature regime peculiarities in brown soils. The fluctuation of temperature in different periods of the year causing positive or negative actions is considered.**

The brown soils are distributed in dried subtropicals of Eastern Georgia on 400-900m above sea level. The influence of temperature on the process of soil formation is great.

The changes in evaporated damp quantity, water transpiration by plants, activity of soil micro and macro fauna were caused by air temperatures.

The experimental plot was chosen in western part of arid zones, exactly in study-testing territory of Agrarian university in Digomi district.

For the observation five soil pits have been prepared (two control virgin soils, the soil of windprotective belts, vineyard and alfalfa soils).

The structure of mentioned soils is as follows: A_1-AB-B_1-B_2-BC_1-BC_2-CD. The surface of each kind of pits is characterized by a good structure that is worsening with the increasing of deepness.

According to the mechanical compositions they belong to heavy clay soils. The actual reaction of soil is often neutral or weak alkaline. In the soils of windprotective belts the reaction is weak acid or neutral. In various soils humus content differed and fluctuated within 3-3.8%. Nitrogen content is rather low too. Phosphorous content is also poor and mainly characterized by biogenes accumulation. Potassium unlike phosphorus is in great amount. The base saturation is heated in absorbed-cations, calcium exceeds (78-88%). The number of carbonates and distribution in each pit has common characteristics, particularly: absence of the above mentioned contents in upper horizons and their appearance together with the increasing of deepness.

Common poriness is rather high (45-60%) in soils. The soil volumetric weight fluctuates within 1-1.5 g/cm^3.

In temperature dynamics two chief periods may be separated: increasing and decreasing. Period of increasing begins in the middle of February or at the beginning of March and lasts till late August, but that of decreasing from the end of August till early February. Those periods were connected with atmosphere (air) temperature regime.

In surface horizons the highest temperature was marked in mid August. In virgin soils it was 19-20^{0}C, under windprotective belts 16-17^{0}C, but in vineyard soils 17-18.5^{0}C and in the alfalfa 17.5-19.5^{0}C.

The data were different enough and relatively low in the soil of windprotective belts. The reason of these differences is a ground litter existence that covers the soil surface by thin layers in windprotective belts. High temperatures were recorded on unplanted virgin lands. Here, soil surface horizons took the heat from superficial air without hindrance. The soils under alfalfa and vineyards took neutral position. Here, those plantations partly hindered heat to reach the upper soil horizons. As for the lower parts of horizons maximum temperature was recorded in mid September. Generally soil is characterized by poor heat conductivity, so in early September temperature went on increasing in lower horizons, while in upper ones it started to decrease (this phenomena is connected with rain that is habitual for Eastern Georgia at that period). In September when temperature reaches its maximum in the upper surface horizons it actively started to decrease.

In winter, mainly in January (sometimes in February) minimal temperature was marked in soils. At that time the soil of windprotective belts was frozen up to 15 cm depth, 18 cm depth in virgin soil, 17 cm depth in vineyard and meadow alfalfa soils. According to this there appeared differences in temperature. The lack of minimal temperature in the soil of windprotective belts in comparison with the other soil temperatures were caused by dead soil cover.

In spring temperature rise in virgin, vineyard and alfalfa soils was much more rapid than in windprotective belts. The process of snowmelt was faster in above-mentioned areas than in the soil of windprotective belts. Biologically active temperatures (>10^{0}C) were mainly recorded since May till October (rarely till November).

In summer period the temperature rise takes place in the whole profiles, while intensive rise and evaporation of humidity relatively cause the drying of upper layers. In such condition plants together with their crown and dead cover protect the upper layers from high temperatures and direct evaporation.

In conclusion, the experimental results revealed that the soils of windprotective belts are relatively characterized by lower temperatures in summer and are less frozen in depth than meadow soils in winter.

Georgian Agrarian University

REFERENCES

1. *T.Urushadze*. Sakartvelos Mta-Tkis Niadagebi. Tbilisi, 1977, 170-180 (Georgian)
2. *G.Talakhadze, K.Mindeli*. Kerdzo Niadagmtsodneoba. Tbilisi, 1976, 223-225 (Georgian).
3. *T.Urushadze, J.Lomidze*. Pochvovedenie. 6, 1977, 55-62 (Russian).

ᲡᲐᲥᲐᲠᲗᲕᲔᲚᲝᲡ ᲛᲔᲪᲜᲘᲔᲠᲔᲑᲐᲗᲐ ᲐᲙᲐᲓᲔᲛᲘᲘᲡ ᲛᲝᲐᲛᲑᲔ, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

M.Gachechiladze, L.Kobakhidze

The Formation of Generative Organs and Character of Pollination of *Trigonocaryum involucratum* (Stev.) Kusn. (*Boraginaceae*)

Presented by Corr. Member of the Academy G.Nakhutsrishvili, August 22, 1996

ABSTRACT. **We studied successive stage of formation of gametophytes, character of pollination and fertilization in embryologically unknown only representatives of genus *Trigonocaryum* - *T.involucratum* (2n = 14). Sex pollination and heterostyle were discovered in species. In short-styled flowers self-pollination was observed (kleistogamy, autogamy), in long-styled ones - cross pollination.**

The numerous family of *Boraginaceae Juss.* embryologically is poorly studied. The family consists of 115 genus and 2500 species [1], only some of the representatives of 24 genus are studied. General data of the embryology of this family is given in the work of Schnarf [3].

For investigation we chose embryologically unstudied endemic genus of Central and Western parts of the Caucasus - *Triginocaryum Trautv.*, represented by single species - *Trigonocaryum involucratum* (Stev.) Kusn. We have studied the successive stages of the generative organs formation and character of pollination.

The research material was taken in Kazbegi region, village Pansheti, at 1800-1900 m above sea level. The plant grows on landslides and stone places. *T.involucratum* is spread in Subalpine and Alpine belts, it blossoms in V-VII month and sets fruits in VII-VIII month [4]. *T.involucratum* is annual, herbaceus plant. Its small-size campanulate flower calyx is 5-dental. On one and the same specimen the corolla is pink, blue and pink-blue. 5 anthers are hidden in the corolla's tube and adhere to it. The ovary is superior, consists of two carpels, divided into 4 lobes by false partition. In each lobe a separate ovule is developed, but only 1-2 coccus reach the final stage of development. The style of the ovary situation is lower than anthers, but sometimes higher. So, in the studied species we have heterostyle. The size of buds is increasing together with their growth and development from 1.5 mm to 3 mm. In them we meet formation of following stages of male and female gametophyte and the fertilized embryo sacs. The size of opening flower is 3-4 mm the endosperm and the embryos in the stage of cotyledons formation are observed. The flowers on the plant are open acropetally.

The formation of generative organs of *T.involucratum* happens in spring.

The anthers are 4-lobed, connected in pairs. Firstly the anther wall consists of epidermis, endotecium, middle ephemeral layer and tapetum. The latter is secretory-type. Cells consist of one nucleus and stay the same till the end. On the two-cell phase of pollen grain the cells of middle layer in anther are already pressed and become flat; lyssis of tapetal cells also occures.

Thus, the wall of mature anther consists of cuticular epidermis and fibrous endotecium.

The sporogenous tissue is represented by angled and thin membranous cells. They are distinguished with dense and abundant cytoplasm.

At the beginning of meiosis in this cells the cytoplasm comes out from cell wall and takes the dumb-bells shape form.

The meiosis often proceeds normally. Arising tetrads are of simultaneous type. Their position generally is izobilateral and sometimes - tetraedral.

More seldom we meet cases when in some specimens of plant flower anther or in its microsporangium meiosis proceeds asynchronously and in anthers may be noticed all phases of division: from prophase I to telophase II.

Uninucleate pollen has delicate plasma. Firstly the nucleus has a central position in it. Then it moves to the periferial part of cell, divides and gives vegetative and generative cells. These cells are almost equal. At the time of the anthers opening the walls between microsporangium destruct and united cavity arises. Pollen grains of different sizes are observed in anthers. Ready pollen is of dumb-bell shape, 2-celled and filled of starch grains. We meet anthers which contain a few number of pollen, or some microsporangium quite empty. Seldom in anther besides 2-celled pollen grains we see 3-celled one.

In *T.involucratum* the number of chromosomes was not known. In tissues of anther cells with 2n=14 chromosomes were observed.

In each lobe of *T.involucratum's* ovary an ovule is layed one by one. The ovules are anatropous: 10-12 layer, tenuinucelated. Its cells are representated by thin cytoplasm along the wall position. The nucleus accordingly has the wall position. We notice the existence of vascular bundles. On the stage of cotyledons formation the number of integuments layer is diminished to 5-6.

One archesporial cell is paid in the subepidermal layer of the ovule, which directly becomes megaspore mother cell. Meiosis goes normally in it and the linear tetrade is formed.

The embryo sac of *T.involucratum* develops from chalazal megaspore according to *Polygonum*-type. Sometimes the degenerated megaspores, which undergo flattening due to growth of lower functional megaspore, are served at the 2 or 4 nuclear stage. The embryo sac is formed in the result of 3 division. The ready embryo sac consists of egg apparatus, the central cell and 3 ephemeral antipodes. The egg cell is located under the pearshaped synergids. Its nucleus is of less size than synergid ones. The pollar nuclei in the central cell are closely arranged. They are fussed before fertilization and form the secondary nucleus of embryo sac. The antipodes are of very little size. They have triangular arrangement.

The meiosis in the male generative sphere happens a little earlier than in the female one. The phase of dyad microspore coincides with anaphase of macrospore; the tetrade phase in microspore corresponds to 2-nuclear stage of embryo sac; at the time of 1-nuclear pollen grain the nuclear stage of embryo sac; at the time of 1-nuclear pollen grain the embryo sac is 4-nuclear, and at the time 2-nuclear pollen grain is 8-nuclear.

Thus, the formation and upening of male and female gametofites happen simultaneously.

Studied species are characterized by sex polymorphism. On some specimens we see bisexual and on some of them female flowers, where the male generative sphere is characterized by several depressive degree. Correspondingly, some anther have few number of pollen and mostly anther or its sporangium is full of sterile pollen or quite empty.

Besides the sex polymorphism the *T.involucratum* flowers are characterized by various styles influencing the pollination process. The most number of individuals are characterized by short styles. In this case it is below the anther. On some specimens we meet flowers with long styles, but the number of such flowers is rather small than that of short-styles.

T.involucratum is basically self-pollinated plant. The pollination mostly takes place in shut flowers kleystogamously. Besides the self-pollination also happens in open flowers, when its style is short and below the anther.

The simultaneous formation of male and female gametophytes give the chance of self-pollination (kleystogamy, autogamy). The pollination in the long-style flowers (where the anther is below the style) generally occurs in cross manner.

After our study it may be concluded that development of female generative sphere of *T.involucratum* in every case proceeds normally. And in the male generative sphere, according to sex change of the flower it develops sometimes normally and sometimes with some disturbances.

In the studied species the self-pollination (kleystogamy, autogamy) may be considered as main form of pollination. The cross pollination takes place less seldom.

The research described in this publication was supported by Grant N KZ 7200 from the International Science Foundation.

N.Ketskhoveli Institute of Botany
Georgian Academy of Sciences

REFERENCES

1. *D.N.Dobrochaeva*. Life of Plants 5(2), Moscow, 1981, 394-398.
2. *V.A.Verechagina*. Family of *Boraginaceae*. Comparative Embryology of Flowering Plants. Leningrad, 1987, 211-216.
3. *K.Schnarf*. Vergleichende Embryologie der Angiospermen. Berlin, 1931.
4. *Z.Gvinianidze*. The Flora of Georgia. **10**, Tbilisi, 1985, 250-251 (Georgian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

HUMAN AND ANIMAL PHYSIOLOGY

D.Lomadze, Z.Nanobashvili, G.Bekaia

On the Peculiarities of Memory in Children with Picnolepsy

Presented by Academician V.Okujava, September 17, 1996

ABSTRACT. **Inattentiveness of patients with picnolepsy is considered to be a general reason for intellectual disorders, displayed by subjects in affective emotional state. Intellectual abilities of subjects reached the normal level as soon as the emotional state became more stable.**

Picnolepsy is known as a form of infant epilepsy. In most of cases prognosis is positive and psychic functions of patients do not differ from those of healthy children as well. On the other hand, epilepsy may cause so-called deviant behaviour, that is well described in scientific publications [1,2]. Enhanced nervous exitation may serve as a reason for deviant behaviour and for altered intellectual abilities of patients as well.

5 patients (7 years old girls) were requested to memorize and recall verbal information (5 words, sentence, short story) given by experimentator orally. The same procedure was repeated again in the case of unsufficient recall.

Inattention of patients with picnolepsy was considered to be a general reason for law scores of subjects in the first series of examination (Table). It appeared rather difficult for them to concentrate on verbal information because of enhanced nervous exitability, displayed in tearfulness, fidgety, dispersion. Even the repetition did not improve the quality of recall.

Table

subjects	recall					
	Initial			After repetition		
	quantity of missed elements					
	words	sentence	story	words	sentence	story
1	2	3	3	3	3	5
2	3	2	2	1	2	3
3	1	1	2	1	2	3
4	3	2	3	4	-	1
5	1	2	3	2	2	4

The scores of patients reached average level established for normal subjects as soon as their emotional state became more stable. Hence, law scores in the first series of examination did not reflect the memory disorders but rather a weakened perception of information, resulted from inattentiveness of subjects.

I. Beritashvili Institute of Physiology
Georgian Academy of Sciences

REFERENCES

1. *N.Kavtaradze.* The Diagnostics and Treatment of Epilepsy. Tbilisi, 1987, 92.
2. *A.I.Boldirev.* The Infant and Juvenile Epilepsy. Moscow, 1990, 316.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

BIOPHYSICS

J.Oniani, V.Yurin, A.Kudriashov,
A.Sokolik, M.Chokhonelidze, V.Demidchik

On the Mechanism of Membranotropic Action of the Heavy Metal Ions on the Plant Cells

Presented by Academician T.Oniani, December 3, 1996

ABSTRACT. **The regularities of interaction between the heavy metal ions and plasmalemma, as evidenced by measurments of the bioelectrical reactions of the cell and cyclosis velocity, are described. It is shown that Pb^{2+}, Zn^{2+}, Cd^{2+}, Ni^{2+} ions in concentration of 10^{-6} – 10^{-3} M possess direct membranotropic effect. The copper in concentration above 10^{-5} M exerts an indirect membranotropic effect in both short – (10–15 min) and long–term (more than 20 min) experiments. In this respect a scheme for the chemical agents' classification is suggested according to the elements of membranotropic effect mechanism and with consideration of the shifts in the ion permeability of membrane.**

The material equilibrium of the plant cell and its environment is regulated mostly by the outer membrane of the cell – plasmalemma. Plasmatic membranes are strongly influenced by the action of chemical agents, which is due either to the direct membranotropic effect or to the indirect disturbance of the biochemical processes in the cell responsible for the maintenance of the normal functioning of the membrane structures [1]. Moreover, some heavy metal ions, on one hand, are the microelements, which are indispensable for normal growth and development of the plant. On the other hand, their presence in the environment, even in a negligible amounts, often causes irreversible changes and death of the plant, which may be due to deterioration of both membrane and intracellular processes [2].

In the above respect it seems expedient to investigate the main rules of interaction between heavy metal ions and plant cell plasmalemma and to classify the observed membranotropic effects in the framework of our approaches [1,3].

MATERIAL AND METHODS

The second – third internodal cells of Nitella were used as an experiment object. Such electrophysiological indices as membrane potential (Ψ), membrane resistance (R), and cytoplasmic motion velocity (V) were recorded.

During recording of the above parameters the cells primary were placed into the solution of Artificial Pond Water (APW), which consisted of 10^{-4} M KCl, 10^{-3} M NaCl, and 10^{-4} M $CaCl_2$ or into the calcium–free colution ARW–Ca (10^{-4}M KCl, 10^{-3}M NaCl). In more details the object, recording methods, and experimental set–up are described in [4,5].

The experiments were carried out according to the following scheme: Control (C)$\rightarrow$C+$Me^{2+}10^{-n}$M$\rightarrow$C+$Me^{2+}10^{-(n-1)}$M$\rightarrow$etc. Afterwards the shift values of the recorded parameters were calculated ($\Delta I=I_{об}-I_{к}$, where I is a recorded parameter, indices "к" and "об" correspond to the cell immersed in the control or in the heavy metal ion containing medium, respectively).

The constant temperature 20°C±2°C and pH (7.2±0.1) were maintained in the solution, which surrounded the cell.

Results and discussion

Out of all the investigated heavy metal ions (Cu^{2+}, Pb^{2+}, Zn^{2+}, Cd^{2+}, Ni^{2+}) the most strong and diverse effect have copper ions. The character of copper action is notably dependent on the time of its action on the cell [6]. Introduction into the solution of calcium ions results in the increase of plasmalemma stability towards heavy metals, which is manifested in decrease of the cellular bioelectric response and the value of threshold concentration in APW versus APW–Ca(Table 1). This effect is due to the well-known membrane stabilizing property of calcium [7].

Table 1

The threshold concentrations, eliciting reliable shift in the bioelectric reactions and cyclosis velocity of the Nitella cells

Ions	Threshold concentration, M		
	Ψ	R	V
APW			
Cu^{2+}	10^{-6}	10^{-5}	10^{-6}
Pb^{2+}	10^{-4}	10^{-3}	10^{-4}
Zn^{2+}	10^{-4}	10^{-4}	10^{-4}
Cd^{2+}	10^{-4}	10^{-6}	10^{-5}
Ni^{2+}	10^{-5}	–	10^{-5}
APW–Ca			
Cu^{2+}	5.10^{-7}	10^{-6}	–
Pb^{2+}	10^{-5}	–	–
Zn^{2+}	5.10^{-5}	10^{-5}	–
Cd^{2+}	10^{-5}	10^{-6}	–
Ni^{2+}	5.10^{-5}	–	–

The ions of Pb^{2+}, Zn^{2+}, Cd^{2+}, in concentration of 10^{-6}–10^{-4}M, elicited membrane depolarization and increase of the resistance as compared with control, regardless of presence or absence of calcium. In a presence of the copper ions the contrary drop of cellular membrane resistance was observed (Fig. 1).

The analysis of experimental data, which characterize shifts of the Ψ and R values, performed in the framework of the model [1, 5], has shown that investigated metals in concentrations of 10^{-5}–10^{-3}M, decrease the permeability of plasmalemma towards K^+ ions (Table 2).

As it was shown earlier [8] absorption of the copper ions from diluted solutions (10^{-5} M) by Nitella cells proceeds slowly. Foolowing 10 minutes of incubation Cu^{2+} are found in the cell walls and in a negligible amounts – in cytoplasm. Along with increase of the copper concentration in the medium to 10^{-4}M, its rapid absorption is obsedved in the cell and after several tenths of minutes the copper is found in the vacuoles. Hence, one can presume that in a concentrations of 10^{-5}M or les and during 10 minute exposition, there is direct interaction between the Cu^{2+} ions and structural components of plasmalemma.

Table 2

Shifts of the membrane ion permeability, induced by the heavy metal ions in the APW.

Ions	Concentration in the medium, M	P	P
Cu^{2+}	5.10^{-7}	+	–
Pb^{2+}	10^{-5}	*	+
Zn^{2+}	10^{-4}	–	+
Cd^{2+}	10^{-3}	–	*
	10^{-4}	–	*
	10^{-3}	–	*
	10^{-6}	+	+
	10^{-5}	–	*
	10^{-4}	–	*
	10^{-3}	–	*

Note: P_k and P_r – membrane permeability coefficient to potassium and other ions: + and – increase and decrease of permeability, respectively; * – uninterpretable shifts.

With an aim to verify this hypothesis experiments were carried out with simultaneous measurement of the cytoplasm motion velocity (cyclosis) and of cell RPD, in a course of action of various concentartions of copper and other heavy metal ions, in short–term (10 min) and more prolonged (30 min) experiments.

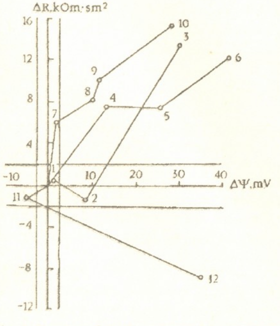


Fig. 1. Results of the heavy metals' testing, as presented on the diagram (ΔΨ, ΔR) lead – 10^{-5}M(1), 10^{-4}M(2), 10^{-3}M(3); zink – 10^{-4}M(4), 10^{-3}M(5), 5.10^{-3}M(6); cadmium 10^{-6}M(7), 10^{-5}M(8), 10^{-4}M(9), 10^{-3}M (10); copper 5.10^{-7}M(11), 10^{-5}M(12).

The heavy metal ions in range of concentration of 10^{-5}–10^{-3}M decreased the cyclosis velocity; nickel (concentration – 10^{-5}M) and copper (concentration – 10^{-7}–$5\cdot10^{-7}$M) were an exception: in above concentration they stimulated cyclosis and elicited membrane hyperpolarization (Fig. 2). Increase of the Nitella cell cytoplasm motion velocity under the influence of 10^{-4} M nickel was also reported earlier [9].

The differences of the cell responses to the action of copper were observed in short and long –lasting experiments. In a long–lasting experiments versus short–lasting ones, the copper ions elict not only more prominent shifts in Ψ and R values, which is in a concordance with earlier publication [6], but they also stronger depress the cyclosis velocity. The other investigated ions did not manifest such differences in a course of

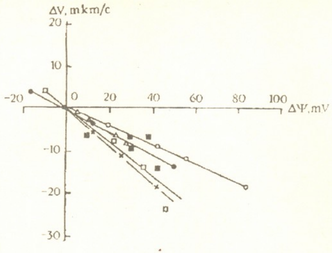


Fig. 2. Comparison of the cyclosis velocity shifts with corresponding potential changes af–ter the action of lead (O—O), zink (X—X), nickel (•—•), cadmium (Δ—Δ) and copper (□) in the short and long–lasting (■) experiments. The points correspond to the different concentrations of the heavy metal ions in the medium.

the experiment time (30 min).

The linear character of potential–dependency ΔV (ΔΨ) of the tested heavy metal ions, which was found in our experiments, confirms the direct membranotropic action [3] of heavy metals in a range of 10^{-6}–10^{-3}M. Deviation from the discovered dependence was found only for copper in a concentration of 2.10^{-5} M, in a short–lasting experiment and during its prolonged action (cf. Fig.2).

We can conclude that the action of investigated heavy metal ions is concerned, in a first place, with the primary changes in a transporting properties of the plasmalemma, e.g. is due to the direct modification of the membrane transport property. The copper ions in a concentration over 10^{-5}M as well as in a range of all the tested concentrations (10^{-7}–10^{-5}M) and in a case of prolonged action probably influence intracellular structures responsible for the cyclosis.

Indeed, in Chlorella cells it was shown that absorption of some heavy metals has a twostage character, namely fast surface reaction independent from the metabolism, and metabolism–dependent slow absorption [10]. The similar regularities are characteristic for the higher plants cells as well [2].

The approaches, which we are elaborating, allow, on the one hand, to reveal direct and indirect mechanisms of the membranotropic action of chemical agents, e.g. heavy metal ions. On the other hand, observed effects could be described in the framework of membrane permeability shifts towards the major potential–changing ions. In this respect it seems possible to classify the chemical agents according to the elements of mechanisms, which underly membranotropic effect, in the framework of the membrane permeability shift, identifying their direct and indirect influences.

I.Javakhishvili Tbilisi State University

Belorussian State University, Minsk

REFERENCES

1. *V.M.Yurin, A.L.Sokolik, A.P.Kudriashov.* Regulation of the Membrane Ion Transport in the Plant Cells. Minsk; 1991, 271 (Russian)
2. *R.M.Welch.* Critical Rev. Plant Sci., **14**, 1995, 49–82.
3. *J.Oniani.* Thesis Doct. Diss. Biol. Sci., Tbilisi, 1987 (Russian).
4. *V.M.Yurin, V.M.Ivanchenko, S.G.Galaktionov.* Regulation of the Plant Cell Membrane Functions. Minsk, 1979, 200 (Russian).

5. *V.M.Yurin.* Bull. Acad. Sci. Belorussian SSR, Biol. Sci.Series, 1, 1989, 24–27 (Belorussian).
6. *V.V.Demidchik, A.I.Sokolik, V.M.Yurin.* Bull. Acad. Sci. Belorussia, Biol. Sci. Series, 3, 1996 (Russian).
7. *V.M.Yurin, M.N.Goncharik, S.G.Galaktionov.* The Ion Transport through the Plant Cells' Membranes. Minsk, 1977, 160 (Russian).
8. *R.Rathsack, H.Sachert.* Ber. Dtsch. Bot. Ges., B. **82**, 1969, 499–504.
9. *R.Y.Jilukas, L.I.Maidunskene* et al. In: Characeae Algae and their Use in the Investigation of the Cellular Biological Processes. Vilnius, 1973, 373–381 (Russian).
10. *D.Cho, S.Lee* et al. Environ. Sci. Health, **29**, 1994, 389–409.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

BIOPHYSICS

V.Yurin, J.Oniani, F.Kudriashov, M.Chokhonelidze, A.Sokolik

Application of Characeae Algae in the System of Ecological Monitoring of the Waters

Presented by Academician T.Oniani, December 3, 1996

ABSTRACT. **The results of the natural waters' testing on the basis of electroalgological analysis are presented. The comparison of the obtained results with the values of the top–permissible concentration of the main pollutants in the waters has shown that with an increase of the pollutant concentration in the assay the recorded cellular membrane parameter shifts are increasing as well.**

In the meantime the biosphere contains about 6 million individual chemical compounds [1], the large part of which is drained into the aquatoria.

Existing system of control, which is based on the evaluation of the Top–Permissible Concentrations (TPC) of the certain substances and other hydrochemical indices, are not sufficient for the full–scale water–protecting effect [2]. The effective means of the integral evaluation of the water quality and of forecasting of ecotoxic danger threatening aquatic flora and fauna, is a bilogical testing method [1,2,3, etc.].

A special importance, in this respect, has development of such devices, which as a major element, contain biological probes. These devices allow to record reactions of the probes and thus accomplish operational and long–lasting control of the changes occuring in the waters. Considering existent standards, on the first stages of biotesting development it is worthwhile to compare the data obtained during biological testing of the natural waters' assayes with those of TPC.

Material and methods

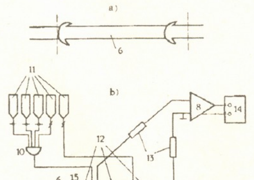


Fig.1. Lay-out of the preparation of the test-object (a) and measurement of electrical properties of the cellular membrane (b): 1,2,3,4 – the chamber compartments; 5 – the chamber; 6 – cell; 7 – current electrodes; 8 – electrometric amplifier; 9 – stimulator; 10 – mixer; 11 – dividing funnels; 12 – isolator (vaseline); 13 – recording electrodes; 14 – automatic recorder; 15 – air–proof lid; 16 – liquid draining hoses.

In the elaboration of the devices for the water sample testing the second–third internodal cells of the Characeae alga Nitella flexilis were used. The algae were grown in the laboratory conditions [4]. The cells were separated from the neighbouring ones and placed into the special bioblock, each chamber of which consisted of 4 compartments (Fig.1). Compartments 1,2 and 3 were filled with control solution; compartment 4 was filled with 10^{-1} M KCl, immitating the intracellular ionic content.

By means of the extracellular electrodes (7 and 13) the electric resting potential differences (RPD) and electrical resistance of the cellular membrane were measured. Then the control solution in the compartment 2 was exchanged for the water sample to be tested and the RPD (ψ) and electrical (R) were registered again, afterwards the changes in the recorded parameters elicited by the test-water sample were evaluated ($\Delta\psi=\psi_{об}-\psi_{к}$; $\Delta R=R_{об}-R_{к}$, where indices "об", and "к" correspond to the cells in test and control samples of liquid, respectively). The method of electroalgologic testing in more detail was described elsewhere [5, 6].

The sampling of the test water assays was performed on the control posts if the two lakes (Lake I and Lake II), situated in a vicinity of the non-ferrous metallurgical plants.

The schedule of assaying was as follows: Control (C) → water sample (dilution – 10^n times) → C → water sample (dilution – 10^{n-1} times) → etc.

As a control sample the background water assay was chosen, which contained a low amount of the evaluated pollutant (F^-, Ni^{2+}, Cu^{2+} and Zn^{2+}). For dilution of the tested water samples the control water samples were used.

The testing of the water samples was carried out in the laboratory conditions at a constant temperature (20°C ± 1°C), pH (7.0 ± 0.2) of the cell surrounding environment and illumination – 300 ± 10 Lux.

Results and discussion

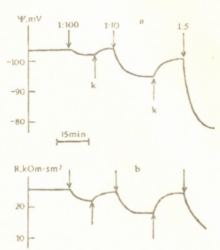


Fig.2 The course of the bioelectric reaction (↓) on the action of water sapmle from the Station 11 and after the wash-out (↑).Shift of the RPD (a) and electrical resistance (b) of the cellular membrane. The figures at the arrows show dilution value.

The presense of most pollutants in concentration equal or exceeding the TPC, elicited considerable changes in the recorded electrical properties of the Nitella cells, when undiluted samples of water were used or even if they were diluted 10–100 times. For example, the water samples from station 20 of the Lake I (near the condensing shop of the mine) contain fluorides and copper, and sample No 11 from the Lake II – nickel, in concentration conssiderably exceeding the TPC. These particular samples, when diluted 100-fold, influenced electrical characteristics of the Nitella cells; The lesser dilution elicited more obvious changes of RPD (15–20 mV) and of cell membrane resistance (10–20 kOhm.cm^2). The example of bioelectrical response development after application of the water sample from St.11 and following wash–out with control solution (water sample from St.6) is shown in Fig.2.

The biotesting results obtained from some water samples of the Lake I and Lake II are given on the diagram $\Delta\psi$, ΔR (Fig.3). The water samples from stations 32, 54, and 54' elicited membrane potential hyperpolarization and increase of resistance versus control. When the water samples from the vicinities of condensing shop (St. 20) were tested, the contrary results were obtained – the difference of electrical potentials and membrane resistance decreased. During testing of the water samples from station 11 of the Lake II the shift of RPD was directed towards depolarization and decrease of cellular membrane electrical resistance, if compared with the background sample of water.

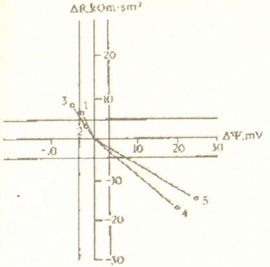


Fig.3 The results of testing of the water samples, presented in diagram ($\Delta\psi$, ΔR): 1 – St.32; 2 – St.54; 3 – St.54'; 4 – St.11; 5 – St.20.

Table 1

Shifts of the membrane ion permeability, induced by the tested water samples

Stations of the assay sampling	P_k	P_r
St.11	*	+
St.32	0	0
St.54	*	–
St.54'	*	–
St.14	0	0
St.20	*	+

Note: 0 no shifts; – decrease; + increase; * uninterpretable shifts

Further exploiting our earlier approaches [7], it seems quite feasible to use shifts of the ψ and R parameters with an aim to classify membranotropic effects, elicited by the tested water samples, according to their action on the ion permeability of the cellular membrane (Table 1). In the simplest case it is possible to differentiate those effects, which are due to the membrane permeability for $K^+ (P_k)$ and other ions P_r [7].

To compare the water samples' electroalgologic testing data with values of TPC let us use the equation [8]:

$$\eta = \frac{C_{Ni}}{TPC_{Ni}} + \frac{C_{Cu}}{TPC_{Cu}} + \frac{C_{Zn}}{TPC_{Zn}} + \frac{C_F}{TPC_F},$$

where C is a concentration of corresponding element.

The comparison of the obtained results show (Fig.1 and Table 2) that sample, which has higher value accordingly elicits higher bioelectric effect. This regularity holds true practically for all the tested water samples.

Table 2

Content of fluorides, copper, zinc, and nickel in the water samples (mg/l) and the values of η

Control stations	F^-	Cu^{2+}	Zn^{2+}	Ni^{2-}	η
St.11	0.46	0.044	0.007	4.16	42.400
St.32	0.30	0.006	0.004	<0.005	0.314
St.54	0.30	0.007	0.008	0.009	0.388
St.54'	0.43	0.02	0.006	0.011	0.603
St.14	0.18	0.009	0.007	0.005	0.267
St.20	21.0	6.000	–	–	74.000

Thus, even simple comparison of the electrophysiological testing data with the TPC values shows feasibility of our approach for integral evaluation of water quality. Moreover, if in the process of farther testing of water samples from the same control stations, corresponding points will appear in other quadrants or there will be serious changes of recorded parameters, it would be quite possible to pressume that there are some quantitative or qualitative changes in the component content in the controlled natural waters.

Our method has several special features, which chould be emphasized once more: High speed–response time after the effector influence is about 20–30 min; high sensitivity–shifts of the recorded parameters were evident even in the cases when the pollutant concentration in the water samples were less than the TPC values, which allows to detect the pollutant on the early stages of environment contamination; The experimental procedure is simple; There is possibility to automatize sampling and evaluation of the biotesting results – electrical sygnal is recorded.

In conclusion it should be noted that more correct interpretation of the results of biological testing of the water quality, in terms of TPC, could be made with an aid of canonic correlation procedure. Albeit, in such a case more material is required for the training sample.

Belorussian State University, Minsk I.Javakhishvili Tbilisi State University

REFERENCES

1. *G.M.Barenboim, A.G.Malenkov.* The Biologically Active Substance. The New Principles of Search. Moscow, 1986, 386 p. (Russian).
2. *A.N.Krainiukova.* Biological Testing in the Protection of Waters Against Pollution. The Methods of Waters' Biological Testing. Chernogolovka, 1988, 4–14 (Russian).
3. *Yu.Gubski, B.B.Dolgo–Saburov, V.V.Kgrapak.* The Chemical Disasters and Ecology. Kiev, 1993, 224 p. (Russian).
4. *L.N.Vorobiov, G.A.Kurella.* Biofizika, 10, 1965, 788 –795. (Russian).
5. *V.M.Yurin, A.V.Plaks, A.P.Kudriashov.* In: The Methods of Waters' Biological Testing. Chernogolovka, 1988, 33–37. (Russian).
6. *V.M.Yurin.* Bull. Acad. Sci. Belorussian SSR, Biol. Sci. Series, 6, 1979, 108–113 (Russian).

7. *V.M.Yurin.* Bull. Acad. Sci. Belorussia SSR, Biol. Sci. Series, 1, 1989, 24–27 (Belorussian).
8. *S.N.Cherkinski.* The Sanitary Conditions of the Waste Water Sewage into the Aquatoria. Moscow, 1962 (Russian).

BIOCHEMISTRY

N.Giabashvili, M.Tsartsidze

Hydrogen Peroxide Production by Cladosporium Herbarum Fungus and Possibility of Its Participation in Degradation of Synthetic Polymers

Presented by Corr. Member of the Academy N.Aleksidse, July 22, 1996

ABSTRACT. **The possibility of hydrogen peroxide production in cladosporium herbarum infected cultural medium was studied. The mechanisms of biodegradation of the synthetic polymers with participation of the hydrogen peroxide were disscused.**

Earlier we have supposed that in addition to the organic acids [1] and enzyme systems [2] the oxidative processes, viz. active forms of oxygen [3,4] play an important role in biodegradation of various materials. However, oxygen intermediates produced by micromycetes are unlikely to reach the culture medium: due to their low reactivity they would be rapidly inactivated as a result of interaction with biosubstrates of the fungus. Consequently it is the active form of oxygen formed in the micromycete infected culture medium that are playing a significant role in degradation of the synthetic polymers. Proceeding from the above said, the goal of our work is to study the mechanisms of the oxidative processes occurring in the culture medium in the course of degradation of synthetic materials caused by the micromycetes.

In experiments the mold fungus of *Cladosporium* herbarum species obtained from the All-Union collection of microorganisms was used. It was grown in the vials on liquid nutrition medium which contained 15.0 g-*m*ole of sucrose: 0.7 g-mole of KH_2PO_4; 0.5 g-mole of KCl; 0.3 g-mole of K_2HPO_4; 0.5 g-mole of $MgSO_4$; 2.0 g-mole of $NaNO_3$; 0.01 g-mole of $FeSO_4$. The film of aromatic polyimide was preliminarly processed with acetone (to remove lipids), sterilized with 70% ethyl alcohol and incubated in mold fungus infected medium in the thermostat at 27°C during 30 days. Degree of biodegradation of the aromatic polyimide was determined by weighing [5]. The amount of hydrogen peroxide and enzyme activity (glucose oxidase, catalase) was defined in the medium infected with moles fungi in accordance with [6,7,8] while the mass of protein was defined by Lowris method [9]. Six days after cultivation we added inhibitors of the oxidizing processes such as ascorbic acid ($12.5 \cdot 10^{-5}$M) and propylgalate ($12.5 \cdot 10^{-4}$M), to the medium as well as a specific inhibitor of the hydroxyl radical mannitol ($12.5 \cdot 10^{-4}$M).

Detailed analysis of the degradation dynamic of aromatic polyimide with *Cladosporium* herbarum demonstrated that the decrease of polymer was observed after 6 days of medium infecting with micromycete, while 30 days later the polymer was degraded by 40% (Fig. 1).

The effect of antioxidants (ascorbic acid, propilgalate) on the above process was studied to find out the role of oxidative processes in biodegradation of synthetic polymers.Both antioxidants proved to have significant inhibiting effect on the polymer biodegradation (49% and 43%, respectively) (Fig. 1). Thus, oxidative as well as enzymatic processes play a significant role in the biodegradation of synthetic polymers.

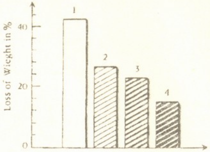


Fig. 1. Changes of biodegradation degree of polyimide film under the effect of antioxidants 30 days after Cultivation of the fungus *CL. herbarum:* 1–control; 2–ascorbic acid; 3–propilgalate; 4–mannitol.

Investigation of the effect of the specific hydroxyl radical inhibitor mannitol on the process of biodegradation of aromatic polyimide demonstrated that mannitol is inhibiting the process by 41% (Fig. 1.). These data alongside with the ones published by us in [3,4] are indicative of significant role of the hydroxyl radical in the oxidative mechanism of degradation of polymer materials. The work of G.K. Skryabin [10] proved hydrogen peroxide serves as a source of hydroxyl radical for P. *chrysosporium* fungus. Our data of using the fungus *Cladosporium herbarum* also testify to the fact that hydroxyl radical like hydrogen peroxide can participate in the degradation of the synthetic polymers and hydrogen peroxide can be considered as one of the sources of the radical.

For this reason of interest was the investigation of the possibility of hydrogen peroxide generation by mold fungus *Cladosporium herbarum*. The dynamics of hydrogen peroxide generation in the process of the fungus cultivation has been studied.It was found that at the initial stage of cultivation (up to 6 days) the culture medium contained the trace amounts of peroxide, 9 days after cultivation this amount significantly increases with gradual subsequent decrease (Table 1).

Table 1

Dynamics of the Enzyme Activity Changes and the Amount of Hydrogen Peroxide (n Mole/min/mg of protein) in the Culture Medium Infected by the *Cladosporium herbarum* Fungus

N	Name of the Index	Time of Cultivation						
		6	9	12	15	18	24	30
1.	Glucose oxidase	0		0		0	0	0
2.	Catalase	–	25		150		185	60
3.	Hydrogen Peroxide	11	40	80	200		101	90

In relation to the above role of hydrogen peroxide in the degradation of polyimide film caused by *Cladosporium herbarum*, the analysis of enzymes glucose oxidase and catalase which generate and degrade the hydrogen peroxide has been carried out.

Glucose oxidase was absent in the fungus infected medium at all stages of cultivation (Table 1). It should be noted that our earlier assumption about the possibility of hydrogen peroxide release by the fungus *Aspergillus niger* [3] was based on the fact of production of glucose oxidase in the culture medium by the above fungus [11].

Interaction of the above enzyme with glucose serves as a source of hydrogen peroxide in culture medium of As. niger. Proceeding from the above said it is evident that production of hydrogen peroxide in *Cladosporium herbarum* is performed with participation of other enzymatic materials.

Determination of catalase activity in cultural liquid shows that it manifests itself in 9 days of the fungus cultivation (Table 1), achieves its peak values on the 15th day (200 nMole/min/mg of protein) with some decrease from 24th day after cultivation (up to 100 nMole/min/mg of protein). It is supposed that release of catalase in the culture medium after 15–20 days of cultivation of the fungus it is necessary to detoxicate high concentrations of hydrogen peroxide which is formed in the culture medium.

Thus,degradation of synthetic polymer by the mold fungus *Cladosporium herbarum* was studied. The possibility of the accumulation of hydroxyl radical and hydrogen peroxide by the fungus infected medium as well as their participation in the process of degradation of synthetic polymers was stated.

I.Javakhishvili Tbilisi State University.

REFERENCES

1. *A.A.Anisimov, A.C.Semjcheva, I.F.Alexandrova, M.S.Feldman.* Urgent Problems of Biodamage. M., 1983, 325.
2. *M.Tien, T.K.Kirk.* Lignin Degrading Enzyme from the Hymenomycete Phanerochaetum chrysosporium Burds. Science, 221, 4511, 1983, 661–669.
3. *M.A.Tsartsidze.* Physical–Abstracts of the Conference of Transcaucasian Republics Microbiologists "Modern Advances in Microbiology and Biotechnology". Tbilisi, 1988, 112–127.
4. *M.L.Koshoridze.* Abstract of Dissertation for Candidate degree, Tbilisi, 1991.
5. *A.M.Keed P.Gilding.* Polymer, 22, 1981, 494–498.
6. *R.K.Root, J.Metcalf, N.Odino, B.Chance.* J. Clin. Invert., **55**, 1975, 945–955.
7. *N.B.Aebi.* Catalase Methods of Enzymatic Analysis. Verlag Chemie, **3**, 1983, 273.
8. *A.H. Chan, D.A.Coffer.* Microbiol. Letter, 15, 1, 1980, 7–12.
9. *O.U.Lowry, N.J.Rosenbrough, A.L.Farr, R.I.Randoll.* J.Biol.Chem. 193,1951, 265–275.
10. *G.K.Skrjabin, L.A.Golovleva, O.V.Maltseva, N.M.Mjasoedov.* Investigation of the Mechanisms of Lignin Degration by P.chrysosporium 1764, M, 3, 1985, 330–338.
11. *J.Marched, L.G.Fracas, J.Austrian, F.Wager.* Appl. Microbiol. and Biotechnol., **30**, 2, 1989, 166–169.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

BIOCHEMISTRY

G.Lekishvili, T.Lekishvili, M.Porchkhidze

Mathematical Investigation of Some of the Biochemical Transformations

Presented by Corr. Member of the Academy N.Aleksidze, July 15, 1996

ABSTRACT. **The use of the new topological indices of the information type is discussed. Some biochemical reactions are studied. The process in connection with the decrease of the information content is substantiated. The directions of the further investigations in this field are outlined. The attempt to ascertain a correlation between structures of molecules and their relative biochemical activities is made.**

Mathematical methods of investigation are widely used in present-day biological and bioorganical chemistry. One should emphasize the model of topological indices based on formal description of molecular structure. In spite of several shortcomings they correlate well with physical and chemical indices of the following type: molecular volume, refraction, affinity to electron, standard entropy etc [1-5]. Especially they are effective at researching the biologically active compounds [6,7]. The requirement of synthesizing a large number of chemical compounds causes increase of the number of factors, which must be envisaged. Mathematical methods allow to exclude the compounds with no essential properties due to very nature of their structure [1].

A graph-discrete mathematical object can be confronted to a molecule. It comprises the finite quantity of vertices and edges, joining some of the vertices. By carrying out the formal operations over a graph, one can confront to the graph a number of the so-called topological indices. For more precise definition of the graphs and the topological indices we recommend to see special literature [2,8].

The topological indices of the information type are of extanding interest. If detaching the X multitude of the selected structural fragments, containing n elements, into K Xi classes of equivalence, so, that $n=ni$, where ni is an amount of the elements of the Xi submultitude, we can express the index of this type via the formula (1):

$$I = 2n \log_2 n - \sum_i n_i \log_2 n_i \qquad (1)$$

The vertices can belong to the identical classes, only in compliance with the following condition:

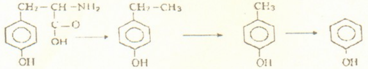
1. The vertices, as well as their neighbours, joined with them, are of the equal "weight" [8].
2. Their neighbours bond the equal number of ^1H.
3. They replace each other after subjecting leastwise one operation of symmerty.

Detaching the molecular graph by the way of this kind the topological indices of so-called theoretical-information type, such as 'IC (Information Content)' can be computed. It is described by the following equation (2)

$$'IC = -\sum \frac{n_i}{n} \log_2 \frac{n_i}{n} \qquad (2)$$

The fact, that the topological indices built this way are named information, is conditioned by their formal likeness with Shannon entropy [9]. These topological indices occupy an expanding place, since the mechanisms of information transfer are observed more often when investigating biological transformations.

For our study we selected the following biochemical transformation:



Tirosine Paraethylphenol Paracresol Phenol

These reactions decsribe one of the ways of forming phenols in living organisms [10].

We computed the above-mentioned topological indices for each substance. It turned out, that in case of the first compound, there existed 11 classes of equivalence ((Fig.1 Table.1). Queue of numerization is free. 'H atoms are not subjected to numerization)

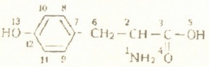


Fig.1.

Table 1

Classes of equivalence	I	II	III	IV	V	VI	VII	VIII	IX	X	XI
Atoms	1	2	3	4	5	6	7	8,9	10,11	12	13

Hence, $'IC_A = 8/13\log_2 13 + 4/13\log_2 6.5 = 3.11$, $I_A = 92.21$. The indices of the molecules were calculated with the same manner:

$'IC_B = 2.72$ $'IC_C = 2.5$ $'IC_D = 1.66$

$'I_B = 53.06$ $I_C = 44$ $I_D = 35.44$

As one can see, the process is carried out with the continuous decrease of information content. It is curious that, in that case, the toxicity of the compounds is connected with the symmetry of their molecules. The higher the symmetry, the higher toxicity is. At the same time, it should be emphasized, that a generalization of that corollary needs further investigation.

So, the topological indices of information type can be used at estimating the likelyhood of biochemical transformations, giving additional possibilities for the investigations of the biochemical reactions taking place in the cell.

I.Javakhishvili Tbilisi State University

REFERENCES

1. *M.I.Stankevich, I.V.Stankevich, N.S.Zephirov*. Topologicheskie Indeksy v Organicheskoi Khimii. Usp. Khim. VII, 1988, 337-366 (Russian).
2. *G.A.Gamsiani, M.H.Gverdtsiteli*. Izomeriis Movlena Matematikuri Kimiis Tvaltakhedvit. Tbilisi. 1992, 74.

3. *M.I.Gverdtsiteli.* Organul Naertta Zogierti Klasis da Mati Gardakmnebis Kvantur-Mekanikuri da Matematikur-Kimiuri Shestsavla. Sadoktoro Disertatsiis Matsne, Tbilisi, 1993 (Georgian).
4. *E.P.Smolenskii.* II Zhurn. Phiz. Khimii, 1964, 38, 1288 (Russian).
5. *N.I.Wiener.* Amer. chem. sos. 69, 17, 1947.
6. *E.Stiuper, U.Briuger, P.Dzhurs.* Mashinnyi Analiz Sviazi Khimicheskoi Struktury i Biologicheskoi Aktivnosti. M., 1982 (Russian).
7. *N.Roshevsky.* Bull. Math. Biophys. 17, 1995, 229.
8. *F.Kharary.* Teoria Grafov. M., 1973 (Russian).
9. *R.A.Gakhokidze, M.I.Gverdtsiteli.* Sak. Metsn. Akad. Matsne (in press). (Georgian).
10. *D.Park.* Metabolism Chuzherodnikh Soedinenii v Zhivikh Organizmakh (Russian).

BIOCHEMISTRY

G.Pruidze, M.Jincharadze, T.Varazishvili, N.Omiadze

Peptides of Black Tea Dry Concentrate and their Aminoacid Composition

Presented by Corr. Member of the Academy D.Jokhadze, July 18, 1996

ABSTRACT. The crude fraction of lomolecular peptides was isolated from black tea concentrate 9 peptides, differing in qualitative and quantitative composition of aminoacids, were identified by paper chromatography.

Peptides play a great role in cell metabolism and plant raw material processing [1,2]. In nature a great number of peptides have high biological activity [3]. A large group of peptides is characterized by antibiotic nature [4], many of them are hormones [5]. Peptides are used in diagnosing of many diseases [6]. Peptides of food product have been recently studied intensively, but tea peptides are rather poorly studied.

The aim of our work is the isolation of peptides from black tea dry concentrates and study of their aminoacid composition.

The isolation was carried out in the following way: tea dry concentrate was dissolved in 70% alchol at continuous stirring for 2 hours.

The received suspence was centrifuged at 1000g, 20 minutes [7]. 10 ml of supernatant liquid was applied to chromatographic paper "Whatman 3MM", as 45 sm long band line. Chromatography was carried out in solvent system: 6-butanol-acetic acid-water 4:1:5. Then 2sm wide band was cut out and developed in ninhydrine reagent. (0.5 gr ninhydrine, 9.5 ml acetone, 1ml icy acetid, 41 ml water), after ninhydrine spraying chromatographic paper was placed in thermostate for 70, 30 min. On chromatographic paper nine pronounced bands of peptides exhibited reaction to ninhydrine. Bands corresponding to separate peptides were cut out from chromatographic paper and peptide elution carried out in 70 % alcohol. Eluates were concentrated in vacuum-rotating evaporator till complete evacuation of alcohol. The received water solution was lyophiecally dried. In order to state peptide aminoacid composition the hydrolysis of peptide preparations was carried out. 4 ml of 6N hydrochloric acid was added to the preparation and it was placed into thermostate at 110c, 24 hours. Hydrochloric acid was evacuated from hydrolisates on water bath by vapourization and the residue dissolved in 10 ml of 20%mm ethylalcohol.

Aminoacid composition was stated by the method of paper chromatography. It was carried out in three fold solvent system: N-butanol-acetic acid-water (8:3:1). Chromatogrames were developed by ninhydrine reaction. Total low molecular peptide fraction was isolated from black tea concentrate. By the method of paper chromatography nine peptides with corresponding Rf-s: 0.242; 0.293; 0.344; 0.422; 0.511; 0.590; 0.662; 0.706; 0.777 were obtained.

Procceding from the table, peptides differ in their aminoacid qualitative and quantitative compositions. Peptides with Rf-s 0.242; 0.293; 0.344; 0.511 contain 6 aminoacids each, but peptides with Rf-s-0.950; 0.662; 0.706 are distinguished by glycine and glutamine. Peptide with Rf-0.590 contain 4 aminoacids, 0.662-5a aminoacids, but peptides with Rf-s 0.706; 0.777-contain 3 aminoacids.

Table 1

Aminoacid composition of black tea dry concentrates

Peptides	Chromatogram Rf	Aminoacid composition
1	0.242	lysin, histidine glycine, glutamic acid
2	0.293	lysin, histidine, aspartic acid glutamic acid, alanine, tyrosine
3	0.344	lysin, histidine, aspartic acid glutamic acid, alanine, tyrosine
4	0.422	lysin, histidine, glycine, aspartic acid glutamic acid treonide, alanine, tyrosine
5	0.511	lysin, histidine, glycine glutamic acid, alanine, tyrosine
6	0.590	glycine, glutamic acid, alanine, tyrosine
7	0.662	glycine, glutamic acid, alanine, proline, tyrosine
8	0.706	glycine, glutamic acid, tyrosine
9	0.777	histidine, glycine, glutamic acid

Thus, 9 peptides were isolated from black tea concentrates, differing in aminoacid qualitative characteristics, that is in particular correlation with Rf-s, determined by paper chromatography.

S.Durmishidse Institute of Plant Biochemistry
Georgian Academy of Sciences

REFERENCES

1. *Y.B.Philippovich*. Uch. 2ap. Mosc. Gos. Pep. Inst. V.I. Lenina 154-160 (Russian).
2. *Nosikova Masaoki*. Tiba Khideo Food Ind. 33, 4, 20-25.
3. Biologically Active Peptides - (Sanaihara noboru) (Kikan, Kagaru Socetsu - 1988, 162-174.
4. Halfsynthetic Peptide Antibiotics (Pat. 4742047 USA, MKI S07 5/12 Oka Masashira Namata Keichi, Konishi Mosataka, N917575, Applic. 10. 10. 88. Publ. 03. 05. 90).
5. *Yakubke, Ch, D., Eshkait Ch,* Anionacids, Peptides, Proteins-Moscow, 1985.
6. Vata Rikuma Sugumu, Katsuki Sigeo/Appl. 58-1783533, Japan, Appl. 1.04.82 N57-54665. Publ. 11.10.83 MKI. S07.c.103156.c. 1201126.
7. *I.A.Asatou, K.G.Ioffe*. Khimia Prirodnikh Soedinenii, 1, 1966, 40-43

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

BIOCHEMISTRY

T.Mitaishvili, Corr. Member of the Academy D.Ugrekhelidze, Ts.Khoshtaria

Nitrobenzene Metabolism in Plants

Presented July 19, 1996

ABSTRACT. **Reduction of nitrogroup into amine group and hydroxylation of aromatic ring of nitrobenzene takes place in maize, pea and kidney bean seedlings. Intermediate products of nitrobenzene metabolism - O-notrophenol and aniline - have been isolated. The studies showed that main metabolic way of nitrobenzene detoxication is conjugation of its intermediates with low-molecular peptides.**

Metabolism of nitrobenzene, as an environmental pollutant has been studied in maize, pea and kidney bean plants. The 1-week-old sterile seedlings under study were placed in water solution of (1-6 ^{14}C) nitrobenzene: specific radioactivity 2.02 10^7 Bq/g, concentration 0.015 Mol/liter, temperature 25^{0}C, exposure time 72 hr. Low-molecular substances were extracted from biomass by boiling with 80% ethanol and divided into three fractions [1]. The labelled substances, after chromatographic separation and autoradiography were identified by spectroscopy and colour test reactions. Fractions of biopolymers were hydrolyzed (6N HCl, 24hr); hydrolyzates alkalified and extracted with diethyl ether The etherous extracts were separated by paper chromatography and labelled compounds were identified according to aforecited mode.

To identify the intermediate products of nitrobenzene metabolism the one-week-old pea seedlings were fed up with water solution of (1-6 ^{14}C)nitrobenzene for 18 hr. After exposure biomass was homogenized in 1N H_2 SO_4 and the homogenate distilled with steam; distillate was alkalified and the unconverted nitrobenzene was separated by steam distillation. Residue was acidified, labelled O-nitrophenol, distilled with steam and identified chromatographically (silufol, solvent system toluene-dioxan 25:4) and spectroscopically (280 nm). (1-6 ^{14}C) nitrobenzene was prepared by nitration of (1-6 ^{14}C)benzene. Evolved $^{14}CO_2$ was absorbed by mixture of monoethanol amine - methyl cellosolve 9:1. The radioactivity was determined by LKB 8100 Liquid Scintillation Counter. Absorption curves were obtained by spectrophotometer Specord UV-VIS; amino acids were determined by analyser AAA-881.

Distribution of radioactivity in plant organs showed that maize, pea and kidney bean seedlings take up nitrobenzene from water solution by roots, and part of labelled carbon is transported to leaves. According to capacity for nitrobenzene absorption plants under study may be disposed in line: kidney bean, pea, maize. Nitrobenzene transport from roots to leaves occured in kidney bean more intensively than in pea and maize. In kidney bean 78% of absorbed by roots nitrobenzene was transported to leaves, while in pea and maize this value was 43.5% and 38% correspondingly.

In the experimental plant tissues (after exposure time 72hr) unchanged nitrobenzene was absent and radioactivity of roots and leaves was conditioned by nitrobenzene conversion products only. Nitrobenzene metabolites were mainly low-molecular, label incorporation into biopolymers was less intensive, and radioactivity of evolved $^{14}CO_2$ was still smaller (Table 1).

The analysis of labelled low-molecular substances showed that they are the peptide conjugates of some nitrobenzene intermediate metabolites. Formation of conjugates in plant cell is possible if only exogenous aromatic substance has an active functional group. Formation of such group in nitrobenzene is possible via reduction of nitro group or via aromatic ring hydroxylation. Aniline, O-nitrophenol and P-nitrophenol (in traces) were isolated and identified from acidic hydrolyzates of conjugates. This points out that both of above mentioned processes proceed in plants under study. It should be noted that nitrobenzene hydroxylation in plants does not follow the rule of orientation and O-nitrophenol is formed instead of meta-isomeride.

Table 1

Distribution of (1-6 ^{14}C) nitrobenzene radioactivity in plants (sp.radioactivity $2.02.10^7$ Bq/g, concentration 0.015 Mol/liter, exposure time 72 hr, 25-27^{0}C).

Plant	Biomass radioactivity, Bq/g	Radioactivity, %		
		Low-molecular substances	Biopolymers	$^{14}CO_2$ per 1g of biomass
Maize				
roots	27.5	84.1	15.9	0.8
leaves	14.4	72.5	27.5	
Kidney bean				
roots	32.3	83.6	16.4	3.6
leaves	27.0	76.7	23.3	
Pea				
roots	33.5	85.7	14.3	1.6
leaves	14.0	65.4	34.6	

The amino acid composition of participating in conjugation low-molecular peptides was identified (Table 2). The same peptide conjugates are formed by metabolism of phenols, benzoic acid, 2,4-dichlorophenoxy acetic acid, and other exogenous substances in plants [2-7].

The intermediates of nitrobenzene transformation in plants - anilin and nitrophenol - after their conjugation with peptides, turn into non-metabolizable form due to functional group blockade. Such substances usually accumulate themselves into vacuoles or are excreted in rhizosphere and therefore belong to the terminal products of metabolism.

Analysis showed that in nutrient solution after experiment no free nitrobenzene was left; solution radioactivity was due to secretion of peptide conjugates by roots.

Short-term experiments (exposure time 18 hr) in pea seedlings showed free nitrobenzene (19% in roots, 38% in leaves), and its free intermediates - aniline and O-nitrophenol, O-nitrophenol amount (free and peptide-conjugated) was in roots 23.1%, and in leaves 5.4 only. Amount of free and bound aniline in roots was equal, but in leaves free aniline was 4.8% and bound - 25.5%. These data suggests, occurence of nitrobenzene hydroxylation mainly in roots, but its conjugation with peptides - in leaves.

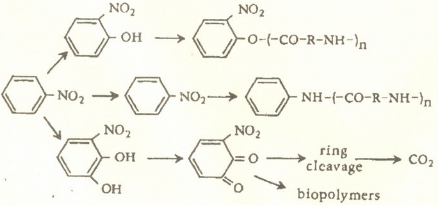
Table 2

Amino acid composition of peptide conjugates of (1-6 ^{14}C) nitrobenzene. (Solvent system: n-butanol - acetic acid - water 4:1:5).

Plant	R_f of conjugates	Amino acid composition
Maize		
roots	0.85	Asp,Glu,Ala,Pro,Tyr,Val.
	0.93	Asp,Glu,Ala,Pro,Tyr,Val.
leaves	0.83	Asp,Glu,Gly,Ala,Leu,Tyr,Val
	0.93	Asp,Gly,Ala,Val,Leu.
Kidney bean		
roots	0.97	Asp.Glu,Ala,Leu,Ser.
leaves	0.84	Asp,Glu,Gly,Ala,Leu.
Pea		
roots	0.95	Glu,Ser,Gly,Ala,Val,Leu.
leaves	0.90	Glu,Ser,Gly,Ala,Val,Leu.

The nitrobenzene label was involved into biopolymer molecules. Amount of labelled biopolymers in leaves was higher than in roots (Table 1). The intermediates of nitrobenzene transformarion were irreversible bound with biopolymers and they are not set free by acid hydrolysis. Structural elements of biopolymers (the sugars and amino acids) had no label. It may be assumed that quinoid substances, formed by oxidative transformation of nitrobenzene metabolites, react with sulfhydryl and amine groups of proteins and irreversible bind with them.

Proceeding from sited above results, the main path ways of nitrobenzene transformation in plants may be presented in the following way:



Georgian Agrarian University.
S.Durmishidze Institute of Plant Biochemistry
Georgian Academy of Sciences

REFERENCES

1. *R.J.Shkolnik, N.G.Doman.* Biokhimija, **25**, 1960,276 (Russian).
2. *D.J.Chrikishvili, D.Sh.Ugrekhelidze, T.I.Mitaishvili.* Bull. Acad.Sci Georgian SSR, **88**, 1, 1977, 173 (Russian).
3. *T.I.Mitaishvili, S.V.Durmishidze, D.Sh.Ugrekhelidze, D.J.Chrikishvili.* Dokladi Akademii Nauk SSSR, **242**,2, 1978, 457 (Russian).
4. *T.I.Mitaishvili, Kh.A.Kakhniashvili, D.Sh.Ugrekhelidze.* In: Metabolism Biosphere Chemical Polutants in Plants. Tbilisi, 1979, 73 (Russian).
5. *D.I.Chkanikov, A.M.Makeev, N.N.Pavlova, T.A.Nazarova.* Phiziologia Rastenii, **29**, 3, 1982, 542 (Russian).
6. *D.Sh.Ugrekhelidze, B.A.Arziani, T.I.Mitaishvili.* Phiziologia Rastenii, **30**, 1, 1983, 102 (Russian).
7. *N.J.Phantskhava, T.I.Mitaishvili, D.Sh.Ugrekhelidze.* Phiziologia Rastenii, **38**, 2, 1991, 386 (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

BIOCHEMISTRY

M.Abutidze, G.Pruidze, N.Mchedlishvili

Purification and Characterization of β-Glucosidase in Tea Leaves

Presented by Corr. Member of the Academy N.Nutsubidze, June 27, 1996

ABSTRACT. β-glucosidase isolated from tea leaves was purified by ammonium sulphate, by ultrafiltration and ion-exchange chromatography on DEAE cellulose. Purification rate is 11. It was shown that enzyme was active at pH 6-6.5 37^{0}C. The dependence of reaction rate of enzyme on the concentration of p-nitrophenyl-β-D-glucopiranoside is governed by Michaelis-Menthen equation.

The crude preparation of β glucosidase was isolated from tea leaves [1] and some of its features were investigated. The aim of this work was the purification of β-glucosidase crude preparation and characteristics of purified enzyme.

The acetone preparation [2] was obtained from tea leaves. Two grammes of acetone preparation was homogenized in glycine buffer, pH 10.2. Obtained mixture was centrifuged during 30 min (3000g). Protein presented in supernatants was fractionated with ammonium sulphate. Precipitates were removed by centrifugation at 10000 g for 10 min. Most of the enzyme precipitated at between 30-80% saturation. This precipitate was treated twice by 0.01 M K-phosphate buffer, pH 6.7. The suspentions were centrifuged at 3000 g for 30 min. Obtained supernatants were united and dialyzed against 0.01 M Na-phosphate buffer, pH 6.5.

This enzyme solution was applied to DEAE cellulose column (2 x 20 cm) equilibrated with 0.01 M Na-phosphate buffer, pH 6.5. The elution of protein was performed with a linear gradient of NaCl from 0 to 1 M in Na-phosphate buffer.

The protein of the enzyme preparation was determined by the method of Lowry et al. [3]. The amount of protein present in chromatography fractions was measured by absorbance at 280 nm.

One unit of β-glucosidase activity was difined as the amount of enzyme catalyzing the realise of 1 nM p-nitrophenol from p-nitrophenyl-β-D glucopyranoside per min..

The profile of elution of β-glucosidase from DEAE is shown in Fig.1.

The results of β-glucosidase purification are summarized in Table 1.

Table 1

Characteristics of Purification of β-glucosidase

Stages of purification	Total protein, mg	Specific activity	Total activity, nm/1min	Purification rate	%
The crude preparation	1005	4.5	4922	1	100
Protein was functioned by ammonium sulphate (0.6-0.8)	256	19.25	4528	4	92
Ion-exchange chromatography on DEAE cellulose	96	49.15	3718	11	75

Some properties of purified β-glucosidase were investigated. Tea leaves β-glucosidase exhibited maximum activity at pH 6-6.5 and T^0=37^{0}C.

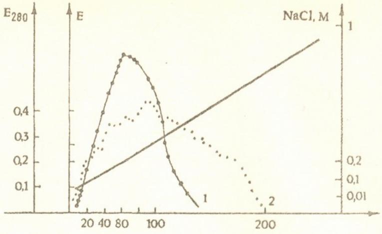


Fig.1. Chromatography of β-glucosidase DEAE-cellulose column:
1 - The activity of β-glucosidase, 2 - The protein.

The dependence of purified β-glucosidase initial rate on the concentration of p-nitrophenyl-β-D glucopyranoside is followed to Michaelis-Menthen equation (Fig.2).

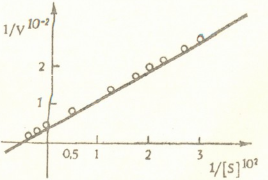


Fig.2. The dependence of β-glucosidase activity on the concentration of n-nitrophenyl-β-D-glucopyranoside.

S.Durmishidze Institute of Plant Biochemistry
Georgian Academy of Sciences

REFERENCES

1. *M.Abutidze, G.Pruidze, K.Gurielidze.* Bull. Georg. Acad. Sci. 147, 3, 1993, 484-486 (Georgian).
2. *G.Pruidze, L.Chachua.* V kn.: Biohimija Rastenij. Tbilisi, 1973, 36-48.
3. *O.H.Lowry, N.J.Rosenbrough, A.L.Farr, R.S.Randall.* J.Biol. Chem., 193, 1, 1951, 265-275.
4. *K.Gurielidze, V.Paseshnichenko, I.Vasilieva.* Dokl. AN SSSR, 3, 1986, 754-757.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

BIOCHEMISTRY

R.Gakhokidze. Z.Khidesheli, L.Tabatadze

On the Problem of the Increase of Plant Resistance to Mycoplasmic Diseases

Presented by Academician G.Kvesitadze, September 13, 1996

ABSTRACT. **The effect of lucreazine on tomato and pepper resistance to the disease of mycoplasmic etiology-stoolbure was studied. Based on the presented results. It can be assumed that after immunification by lucreazine, plant organism assimilates this compound. Plant growth and development are stimulated, a number of its biochemical, biometrical, ultrastructural and anatomical indices improve. The mycoplasmic infection decreases. Such results may be considered as a demonstration of plant increased stability to infections and unfavourable environmental factors.**

Plant mycoplasmic diseases are wide spread and cause considerable damage to agriculture. Mycoplasmic as well as virus diseases, because of specificity of pathogenic microorganisms and infection spreading ways, belong to plant disease with practically impressible cure.Fight against them accentuates mainly on the disease-preventive service [1]. In this aspect seed treatment by chemical immunizators which increases plant general stability to infections and unfavourable environment, is of great importance. In Georgia, the organic-mineral substances – chelate compounds, synthesized at the Institute of Physical and Organic Chemistry of the Georgian Academy of Sciences, are used [2]. To increase plant resistance to mycoplasmic infections, the natural antioxidant-steroids prepared at Moldavian Institute of Ecological Genetics, are recommended in the last years [3]. These substances penetrating into plant cell, fulfill the function of nonspecific resistance inducer, and stimulate defensive mechanisms of plant [4].

The purpose of this work was to investigate the effect of ecologically pure and biologically active substances on various infections for increasing general resistance of agricultural plants. Particular attention was given to vital processes in plants under study and biochemical mechanisms stimulating plant resistance. Laboratory experiments were carried out at the Department of Bioorganic Chemistry of Tbilisi State University and Institute of Plant Protection. Field experiments were carried out in East Georgia (Sarthichala, Varkethili).

We have studied the effect of lucreazine on tomato (var. Peremoga) and pepper (var. Bulgaruli 79) resistance to the disease of mycoplasmic etiology-stool bure. Mixture of salts of microelements (Zn, Cu, Mn, B, Mo) ("Market-Gardener's Hand-Book", 1981) and organic-mineral chelate were taken as a standard, and untreated seeds - as a control. Exposure time was 24 hr. The energy of germination and germinating ability of tomato and pepper seeds have been measured in Jacobson apparatus. Biometric measurements of seedlings have been done. The results are presented in Table 1.

Table 1

Plant	Variant	Energy of germination, %	Germinating ability %	Length of sprout, cm	Length of root, cm
Tomato	Lucreazine	79.2	92.7	2.4	4.6
	Chelate	71.0	85.8	2.1	3.8
	Microelements	65.2	80.8	1.8	3.6
	Control	51.3	77.0	1.7	3.2
Pepper	Lucreazine	65.2	90.6	3.5	7.6
	Chelate	56.3	81.4	2.8	6.8
	Microelements	50.5	78.5	2.6	6.1
	Control	44.3	78.1	2.1	5.9

The biometric measurements of tomato and pepper seedlings grown from the treated seeds were carried out at hot-house of the Institute of Plant Protection. The results are presented in Table 2.

Table 2

Plant	Variant	Length of stem, cm	Length of root, cm	Weight of stem, g	Weight of root, g	Diameter of stem, mm	Total wt of seedling, g
Tomato	Lucreazine	20.5	7.7	5.3	1.0	4.1	6.3
	Chelate	18.9	6.8	4.2	0.8	3.8	5.0
	Microelements	18.1	6.2	4.0	0.5	3.6	4.5
	Control	17.8	6.0	3.9	0.5	3.2	4.4
Pepper	Lucreazine	33.2	7.4	11.8	1.8	4.0	14.6
	Chelate	30.8	5.9	9.6	1.5	3.8	11.1
	Microelements	30.0	5.2	7.7	1.6	3.6	9.3
	Control	26.7	5.0	6.5	1.2	3.2	7.7

Presented in Tables 1 and 2 data indicate that energy of germination and germinating ability of seeds increase due to their treatment with lucreazine; the biometric indices of seedlings also improve considerably. Intensive development of roots, should especially be noted increase of length and diameter of stem.The experimental plants were more deep-green and visually differed from control plants.

Experiments were carried out to study the effect of lucreazine on some biochemical indices of plants, and also on the activity of respiratory ensymes-catalase, peroxidase and polyphenoloxidase. Activity of enzymes was estimated by the Mikhlin and Bronovitskaja method, and ascorbic acid content – iodometrically.

Data presented in Tables 3 and 4 show that as compared with standard and control, in both plants percentage content of dry matter, total sugar and vitamin C, increases and the activity of enzymes catalase, peroxidase and polyphenoloxidase also increase.

Under the influence of mycoplasmic infection a number of pathological alterations takes place in plant ultrastructure. First of all the chloroplast structure is getting destroyed; in particular, thylacoides desintegrate, number and dimensions of osmiophil and lipoid globules and starch granules increase [5]. Moreover in infected cell the autolytic processes are intensified. As a result the biomembrane structure destroys, the substrate content into cell decreases, and the cell dies [6].

Table 3
Effect of lucreazine on some biochemical indices of tomato and pepper fruits.

Plant	Variant	Dry matter, %	Total sugar, %	Vitamin C, mg %
Tomato	Lucreazine	4.9	2.5	30.2
	Chelate	4.4	1.2	23.8
	Control	4.2	1.1	18.2
Pepper	Lucreazine	9.2	3.4	280.5
	Chelate	8.6	2.7	258.0
	Control	8.3	2.1	212.6

Table 4
Effect of lucreazine on the activity of enzymes in tomato and pepper leaves.

Plant	Variant	Catalase ml g 4min.	Peroxidase 0,01 J plant	Polyphenoloxidase 0,01 J plant
Tomato	Lucrezine	564	12.48	5.32
	Chelate	520	11.64	4.46
	Control	496	10.12	4.18
Pepper	Lucrezine	347	9.09	3.86
	Chelate	306	7.95	3.24
	Control	290	7.23	2.98

Biologically active substances partially change the metabolism of infected plant, activate its protective reactions, and as a result stabilize ultrastructure of cell. We have made up our mind to study the ultrastructural changes proceeding in plants treated by lucreazine, by means of electronic microscopy.

a b
Fig. 1. Chloroplasts in tomato tissues, infected with mycoplasma
a. Control. b. Treated with lucreazine.

The method of ultrathin slices has been used for investigations by electronic microscopy. As objects under study the bearers of mycoplasmic symptomes tomato leaves were taken untreated and treated by lucreazine. Material was fixed by 5% glutaric aldehyde solution in cacodylate buffer during 4 hr [7], and postfixed by 1% osmium tetraoxide solution in cacodylate buffer for 2 hr. Material was dehydrated by alcohol of increasing concentrations and absolute acetone, after which it was placed on the ultramicrotome, (Richert) contrasted by uranyl acetate and plumbic citrate [9], and examined by electronic microscope (JEM-1200 EX).

In the cell structure of infected plants a number of pathological alterations were to be observed. These alterations were most visible in chloroplasts and expressed by desintegration of thylacoidal granas and stroma. On the destroyed part of lamellar-granal system the lipophilic and osmiophilic globules were to be observed. Dimensions and number of starch granules strongly increased,and therefore chloroplasts were much degraded. Chloroplasts, deformed and immoderately swollen,occupied the whole space of the cell (Fig. 1a). Analysis of ultrathin slices of treated by lucreazine plants showed that here choroplasts are relatively stable, and in spite of mycoplasmic infection they preserve characteristic lamellar-granal system. Thylacoides were normally developed and organized structurally more clearly, than in control plants.Starch granules were relatively less and smaller (Fig. 1b). Thus, it may be said that after treatment of infected with mycoplasmosis plants by lucreazine, safety of cell and their organelles increases, and ultrastructural organization normalizes.In our opinion such dependence, is connected with protective mechanism of plant. Anatomical measurements of untreated and treated by lucreazine tomato leaves showed that in latter width of epidermis and height of palisade and spongy parenchymas enlarge (Table 5).

Table 5

Variant	Width of epidermis, micron	Height of palisade parenchyma, micron	Height of spongy parenchyma, micron
Lucreazine	5.8	7.2	12.0
Control	5.0	6.7	10.8

Relatively dense structure of palisade parenchyma and width of epidermal cells form more powerful barrier against infection penetration of proboscis by insect-vehicle and spreading of infection.

Field experiments, carried out in 1994-1995 against a background of mycoplasmosis, showed that among plants treated with biologically active substance, the percentage of infectivity decreases by 30-35 per cent against control plants. Mycoplasmic symptoms were visually feebly expressed, and experimental plants maintained longer the vital activity.

Thus, based on the presented results, it can be assumed that after immunification by lucreazine, assimilating this compound in plant organism, plant growth and development are stimulated, a number of its biochemical, biometrical, ultrastructural and anatomical indices improve. The mycoplasmic infection decreases. Such results may be considered as a demonstration of plant increased stability to infections and unfavourable environmental factors.

I.Javakhishvili Tbilisi State University

REFERENCES

1. *Yu. I. Vlasov, A.N. Samsonova, G.D. Kaverzneva.* Phitomycoplasms: Classification, Diagnosis and Means of Struggle. M., 1987.
2. Metal chelates of natural compounds and their application. Tbilisi, 1974.
3. Recommendations for the Protection of Tomatoes from Mycoplasmatic Diseases-Stolbure. Volgograd, 1986.
4. Application of Steroidal Glycosides for the Increase of Total and Antivirous Stability of Vegetable Cultures. Kishinev, 1985.
5. E.*F. Ulanova.* Changes of Cellular Structures of Tomatoes Induced by Mycoplasma. Virous Diseases of Farm Cultures. M., 1980
6. V.*L. Fejutina.* Mycoplasms in Plant Cells. Materials of the All-Union Conference. Leningrad, 1978.
7. *D.D.Sabatini, I. Miller, R.J. Barnett.* J. Histochem. Cytochem., 12, 1, 1964.
8. G.*H. Luft.* J. Biophys. Cytol., **9-5**, 1961.

E.*S. Reynolds.* J. Cell. Biol., **17,** 1963.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

CYTOLOGY

Kh.Janelidze

Ultrastructural Changes in Retinal Photoreceptors in Hens, Hen Embryos and Newly-Hatched Chickens under Ionizing Radiation

Presented by Corr. Member of the Academy B.Kurashvili, August 8, 1996

ABSTRACT. **Ultrastructural changes in retinal photoreceptors have been studied in hens, hen embryos and newly-hatched chickens under the ionizing radiation.**

It has been shown that radiation affects the external segments of both rods and cones, causing dezorientation, fragmentation and disappearance of their membrane discs. Changes take place in the internal segments of photoreceptors as well: mitochondrial double membranes and cristas disappear in the ellipsoid, while in the paraboloid the disintegration of glycogen granules and a slight decrease in their number are observed. In the cytoplasm of visual cells polysome complexes break down into separate ribosomes and a slight decrease in the diameter and length of the endoplasmic net elements is noted. In hen embryoes of 12 days the radiation delays the formation of membrane discs.

It has been supposed that under radiation redox processes decrease in photoreceptor cells and the level of protein synthesis also comes down, causing delay in the formation of membrane discs.

A ray-sensitive part of the eye retina presents photoreceptors with their rather complex structure. Ultrastructural ingredients of photoreceptors are at the same time their functional units which differ by both structure and function. It is known that the external segment of photoreceptors consists of membrane discs taking part in the light perception. The molecular base of membrane discs is made of protein opsin and retinol (the oxygenated form of A-vitamin) which in complex present a pigment of vision-radopsin [1]. Under the action of **phaton** (the energy of light) the chemical relation between opsin and retinal breaks off, causing the irritation of the visual cell [1].

The internal segment consists of an ellipsoid which presents mitochondrial clusters. It provides energy to the processes of irritation and ensures the vital energy of the cell [1]. The endoplasmic net and paraboloid ensure the protein synthesis and provide the cell with glycogen - a source of energy. The body of photoreceptors consists of a nucleus and cytoplasm. It is connected to the appendages of bipolar and horizontal cells through the synapses [1].

The bird's eye retina consists of two photoreceptors - the rod and cone which ensures black and white and coloured vision.

Earlier it was shown that 4 hours' radiation calls forth some changes in definite character in the rabbit's retina, particularly: in the outer nuclear layer of photoreceptors the nuclei are picnotized, the outer basal membrane is fragmented; the cells of the inner nuclear layer [2] are also picnotized.

The changes in structural components of photoreceptors under the ionizing radiation seem not to be sufficiently studied. This induced us to examine the problem more deeply. Our object was to study ultrastructural changes in photoreceptor cells in birds (hens in particular) under the action of the ionizing radiation on the eye. The eye retina of adult hens, newly-hatched chickens and hen embryos of 12 days was studied. The studied stuff was treated by a wellknown electron microscopical method. The stuff was fixed in the 4-oxygen osmium solution, in epon 812; the ultrathin sections were prepared on the Reikhard-type Ultramicrotome and studied with the Japanese Electron Microscope Y-100 B. The adult hens' eyes were *irradiated. with 4600 X*-rays, 299 kv, *15 mA* during 10 min. The newly-hatched chickens' eyes were treated with 2300 X-rays, 200 kv, 15 mA during 5 min. The hen embryos of 12 days were treated with 2300 X-rays, 200 kv, 15 mA during 1 min and 30 sec. These doses of the ionizing radiation were specially used in order to exclude a lethal action of the radiation.

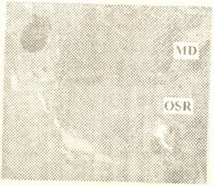


Fig.1. Fragment of adult intact hen's visual cell. OSR – outer segment of rod; MD – membrane discs. x20000; in epon-812

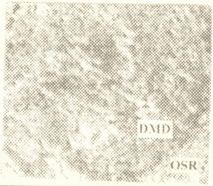


Fig.2. Fragment of adult hen's visual cell after the ionizing radiation. OSR – outer segment of rods; DMD – disintegrated membrane discs. x25000 in epon-812

Fig.3. Fragment of adult intact hen's visual cell. OSC – outer segment of cone, DF – drop of fat, x25000, in epon-812

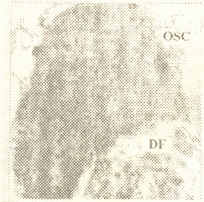


Fig.4. Fragment of adult hen's visual cell after the ionizing radiation, OSC – outer segment of cone, DM – disintegrated membranes, DF – drop of fat, x25999, in epon-812

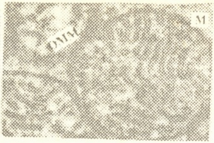


Fig.5. Ellipsoid of adult intact hen's rod. M – mitochondria; OMM – outer double membrane of mitochondrion; x48000, in epon-812

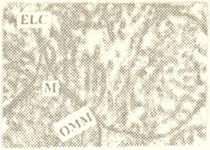


Fig.6. Ellipsoid of adult intact hen's cone. ELC – ellipsoid of cone; M – mitochondrion; OMM – outer double membrane of mitochondrion, x45000 in epon-812

The results obtained show that under radiation ultrastructural changes in definite character take place in the photoreceptors of adult hens and newly-hatched chickens. These changes consist in the following: in the outer segment of photoreceptors (both rods and cones) the strictly regulated orientation of membranes is disturbed in

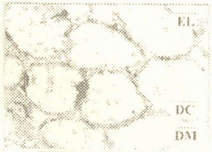


Fig.7. Ellipsoid of adult hen's visual cell (rod) after the ionizing radiation. EL – ellipsoid; DM – disintegrated mitochondrion; DC – disintegrated cristas, x30000; in epon-812

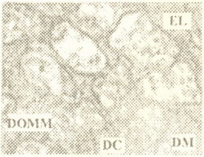


Fig.8. Ellipsoid of adult hen's visual cell (cone) after the ionizing radiation. EL – ellipsoid; DM – disintegrated mitochondria; DC – disintegrated cristas; DOMM – disintegrated outer membrane of mitochondria; x25000; in epon-812

comparison with standard (Fig.1,2,3,4). Ultrastructural changes are observed in the second structural component of photoreceptors, i.e. in the ellipsoid of the inner segment which presents mitochondrial clusters and is placed in rods directly under the outer segment, while in cones - under the fat drop. In the norm mitochondria have a well expressed double membrane, and the cristas are placed across the longitudinal axis of mitochondria (Fig.5,6). Under the ionizing radiation the double membrane and cristas get disintegrated and the matrix density reduced against the norm in the mitochondria of the adult hens ellipsoid (Fig.7,8). In the photoreceptor cytoplasm

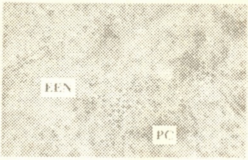


Fig.9. Fragment of adult hen's intact visual cell. EEN – elements of endoplasmatic net; PC – polyribosome complexes; x20000, in epon-812

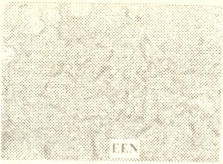


Fig.10. Fragment of adult hen's visual cell after the ionizing radiation. EEN – elements of endoplasmatic net; x20000; in epon-812

polysome complexes break up into separate ribosomes, the elements of endoplasmic net become degranulated and their diameter decreases against the norm (Fig.9,10). The glycogen grains of paraboloid break up and their size and number are reduced in comparison with the norm (Fig.11,12). In the photoreceptor nucleus chromosome grains unite into clusters and settle basically near the double membrane of the nucleus.

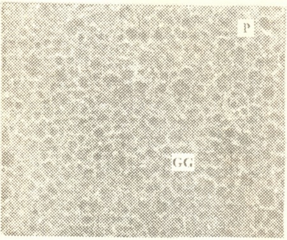


Fig.11. Fragment of adult intact hen's visual cell. P – paraboloid; GG – glycogen grains; x60000 in epon-812

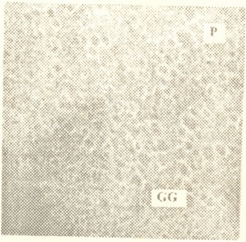


Fig.12. Fragment of adult hen's visual cell after the ionizing radiation. P – paraboloid; GG – glycogen grains; x60000 in epon-812

Under the effect of the ionizing radiation analogous ultrastructural changes are observed in the photoreceptors of newly-hatched chickens. Particularly, the dezorientation and fragmentation of membrane discs take place in the outer segment of visual cells against the norm. In the ellipsoid the swelling of mitochondria and fragmentation of cristas and their partial disintegration are observed. In the photoreceptor cytoplasm polysome complexes break down into separate ribosomes.

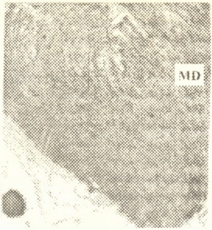


Fig.13. Fragment of visual cell of intact hen's embryo (of 14 days); MD – membrane discs; x45000 in epon-812

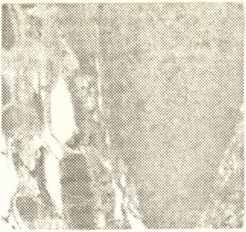


Fig.14. Fragment of visual cell of hen's embryo (of 14 days) under radiation. x40000 in epon-812

The formation of the first membrane discs is known to take place in the outer segment of photoreceptors in hens' embryoes on the 12th day of the embrional development [3]. On the following days the number of membrane discs increases. On the 14th day of the embryonal development a rather large number of membrane discs is observed in the outer segment of photoreceptors (Fig.13). 2 days after the radiation, i.e. on the 14th day of the embryonal development the membrane formation stops in the outer segment of photoreceptors of hens' embryos (of 12 days) (Fig.14).

The work done allows to conclude that in adult hens and newly-hatched chickens the ionizing radiation causes strong degenerative changes in visual cells (both rods and cones) expressed by the dezorientation and break-down of membrane discs in the outer segment, and stop of the membrane discs formation in hens' embryos. Pathologic changes are also observed in the second structural component of photoreceptors, i.e. in the ellipsoid of the inner segment: the mitochondrial double membranes and cristas break down, causing decrease in the redox potential.The energy of redox processes can be used in the synthesis processes fulfilled by the photoreceptors, particularly in the protein opsin synthesis which constantly goes on in visual cells and is needed for reducing membrane discs.

In our case the degranulation of the endoplasmic net elements and disintegration of polysome complexes into separate ribosomes in the cell cytoplasm points to a low level of protein synthesis in photoreceptors. The glycogen grains disintegration in the paraboloid of photoreceptors also points to a low energetic level in the photoreceptors. A drop in the quantity of hydrocarbons was noted by Frens who, using the hystological method, showed a decrease of succinate dehydrogenase and glycogen in the rabbit's retina under the effect of radiation of 6000 X-rays on the eye during different lengths of time [5].

The study showed the halting of membrane discs formation in the outer segment of photoreceptors in hens' embryos (of 12 days) on the 14th day of the embrional development, i.e. 2 days after the radiation. We put it down to a low energetic level of metabolism in the photoreceptor cells, unable to provide the cell with ingredients; the latters are necessary for the formation of ultrastructural components of photoreceptors. The synthesis of protein opsin is disturbed and the membrane discs are deprived of this protein. Protein opsin and retinal (the oxygenated form of A-vitanin) are known to make a molecular base of membrane discs [1]. The continuous synthesis of opsin and its transportative from the cytoplasm to the membrane discs of the outer segment of rods is well known in Young's and Droz's works of electrone-microscopic character where they used labelled aminoacids [4].

Institute of Zoology
Georgian Academy of Sciences

REFERENCES

1. *I.A.Vinnikov.* Tsitologicheskie i Molekuliarnie Osnovi Retseptsii. L., 1971.
2. *P.Sidney, Kent and A.Arnold.* Influence of High Intensity of X-Radiation on Retina: Hystological, Histochemical and Chemical Research of Retina, 1970. Division of pathology, the School of Airforce Medicine USAF, Randolph Air Force Base, Texas.
3. *Kh.N.Janelidze, G.S.Kvinikhidze.* J. Tsitologia, **XXIII**, 1, 1981.
4. *C.M.Frances.* J.Physiol., **119**, 1953.
5. *R.W.Young, B.Droz.* J.Cell.Biol., **39**, 1, 1968.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

CYTOLOGY

R.Kapanadze, L.Ninoshvili

Autoimmune Myocardiopathy: Functional, Structural and Biochemical Changes

Presented by Corr. Member of the Academy T.Dekanosidze, August 29, 1996

ABSTRACT. **Experimental researches were carried out on male rats. Autoimmune cardiomyopathy resulted from adjuvant Freudni. Heavy changes of structural, ultrastructural, biochemical and functional type were traced down in myocardia. Reomorphology of blood was subjected to changes too. The latter is of great importance for systolic function of myocardia. In order to correct broken metabolism of myocardia and α-Tocopherol and Cytochrome C are effectively used.**

In literature there have been reported complete investigations based on the study of structural, ultrastructural, functional, metabolic and biochemical changes in cardiomyocytes [1,2,3,4].

The goal of the present investigations was to study some functional, hystochemical, biochemical changes occurring in cardiomyocytes in autoimmune cardiomyopathy (AICM). It should be also pointed out that the state of lipid metabolism and lipid peroxidation fermental activity in AICM is not quite clear.

Recent investigations indicate the role of disturbances in lipid metabolism, lipid peroxidation processes in different pathologies including cardiovascular system with decreased myocardial contractility.

Experiments were carried out on 95 male rats (180-200 g of weight). AICM was induced by abdominal subcutaneous injection of a mixture of adjuvant Freudi 0.5cm^3 and 0.5cm^3 heart homogenate. Secondary immunization was achieved by injection to the same animals of 0.25cm^3 adjuvant Freudi +0.25cm^3 heart homogenate one month after first immunization. ECG in 3 standard leads was taken on the 20-30th day after first immunization and before slaughter of the animals.

30 animals were treated for 10 days with α-tocopherol (10 animals), cytochrome C (10 animals) and cytochrome C + α-tocopherol (10 animals). Blood samples for hemomorphological and biochemical methods were drawn from the tail vein.

To study myocardial structural and metabolic changes general morphological, histoenzymochemical methods were used together with biochemical methods: electron microscopy and rustral electron microscopy for the study of blood rheology.

As a result of such complex study it has been established that in AICM the decreased contactile function of the heart was determined by disturbances in metabolic processes, structural and ultrastructural changes in the heart muscle. On the third month after induction of AICM ventricular premature beats and ventricular tachicardia were recorded on ECG.

Macroscopically the heart was sluggish and balllike. Its mass increased up to 1.48 gram (normally 0.9 g in average). General morphological and hystochemical methods have revealed myocardial stromal and parenchymal damage. Stromal hemodynamic changes were characterized by accumulation of oedemal fluid between cardiomyocytes and small vessels. Intermedial tissue was rich on gentle connective tissue and argilofill fibres; there was also inflammatory infiltration consisting of lymphoid and histocyte cells. In myocardial parenchyma uneven hypertrophy of cardiomyocytes, their granular distrophy, lysis and homogenization have been observed.

Hypertrophied cardiomyocytes are rich on large granules of glycogen rather than on fine granules. This indicates to decreased protein-bound glycogen content and possible unstability of metabolism. In AICM the content of RNA is decreased in hypertrophic as well as in nonhypertrophic cardiomyocytes. Strongly positive reaction for RNA was in cytoplasm of those cardiomyocytes which were in the state of lysis indicating to the discomplexion of RNA in them. In some groups cytoplasm activation of enzimes, taking part in oxidation processes was decreased (succinate-hydrogenase, NAD and NADP diaphorases, cytochrome oxidase) while adenosentriphosphatase activity in some hypertrophic myocytes was unevenly increased. Acid phosphatase activation is also increased while that of alkal phosphatase decreased.

Electromicroscopically in cardiomyocytes besides the changed ultrastructural elements of cells, cells with preserved function were found. Cardiomyocytes are enlarged with mitochondrial hyperplasia and myofibrillar hypertrophy.

Very often in hypertrophied cardiomyocytes intact ultrastructural elements have been observed. Such cardiomyocytes execute the function of damaged myocytes being in the state of parenchymal dystrophy.

In AICM total myocardial phosphorus content of phospholipids decreases by 30 per cent.

Among the phospholipids fractions phosphadilserin, sphingomyelin, phosphatilcholine, cardiolipine content was especially low, neutral and acid phospholipid ratio was increased. This indicates to disturbances in phospholipids synthesis. Increase in malondialdehyd terminal products content and catalase activation in erythrocytes and serum were also observed. Thus biochemical studies have shown that in AICM myocardial metabolic disturbances occur. They are expressed by enhanced lipids peroxidation, changed phospholipids spectrum, increased activation in antioxidative enzymes.

Rheological study of the rat's blood revealed the changed architectonics and form of the surface of the erythrocytes, changed to worse their deformation ability, increased hematocrit, increased blood viscosity and decreased catecholamines content. All the above mentioned draws to conclusion that in AICM morphofunctional and hemorheological changes in erythrocytes are in close relationship and interact with each other. Rheological disturbances are secondary to morphological changes in erythrocytes. This can contribute to better diagnosis, treatment and more correct evaluation of the pathological process.

The treatment of AICM with α-tocopherol for 10 days revealed the decrease in myocardial malondialdehyde content and the trend to the improvement in phospholipids spectrum regulation. Springomyelin and cardiolipine content in cardiomyocytes increases, mitochondrial membrane structures are being normalized and their functional activity is enhanced. All this is in close correlation with myocardial sctructural and enzymochemical activity.

As a result of treatment of rats with cytochrome C heart rhythm disturbances were met significantly less frequently, T wave returned to normal, glycogen and RNA content in cardiomyocytes increased as well as the activity of enzymes taking part in restorative-oxidative and hydrolysis processes, the ultrastructural changes became less intensive and hemorheological parameters were regulated.

The treatment of animals with cytochrome C + α-tocopherol turned out to be even more effective. Such combination enhanced structural and ultrastructural restoration rate and improved myocardial contractile ability.

M.Tsinamdzgvrishvili Clinical and
Experimental Institute of Cardiology

REFERENCES

1. *A.V.Kainova, I.B.Kuraeva.* J. Exper. Biol. & Med. **3**, 1961, 60-64 (Russian).
2. *E.V.Burlakova.* In the book "Biochemistry of lipids and their role in metabolism", M., 1981, 13-15 (Russian).
3. *N.N.Kipshidze, I.A.Kapba.* Cardiologia, **10**, 1983, 13-15 (Russian).
4. *S.I.Kostin.* Archives of Pathology, **11**, 1989, 47-53 (Russian).

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

PALEOBIOLOGY

V.Chkhikvadze, N.Amiranashvili

History of the Fossil Turtles Studies in the Caucasus

Presented by Academician L.Gabunia, July 8, 1996

ABSTRACT. **The history of investigation of the fossil turtles of the Caucasus, region by region, is reviewed. The list of the extinct species is presented as well as their stratigraphical and geological distribution.**

It is high time that paleoherpetological research in the Caucasus should be summarized, seeing that an ever increasing inflow of new information has almost eclipsed the work of earlier investigators that forms the present day framework of scientific knowledge. The majority of earlier publications appeared in numerous sources some of which have now become rarities. More often these works are sitting inconspicuously on the lists of references appended to reviews, while some of them have never been quoted even by contemporary fellow researchers. This inspired us to write a history of studies in the herpetofauna of the Caucasus, with a view of summarizing results and outlining the prospects of further investigations. Unlike the other similar publications, our material is presented not in the chronological order, but rather under separate headings.

The fossil herpetofauna of the Caucasus has been studied disproportionately. Yet turtles have been covered almost to the full. Therefore we shall confine our work to outlining the history of studying of these particular reptiles.

Georgia. The first mention of the fossil remnants of turtles found in Transcaucasia was made by V.V.Bogachov [1], who says in his article that "from Mirsaani (Shirak District) to Eldar in Southern Kakheti, variegated clay deposits are quite abundant, which have preserved remnants of hipparion, a rhinoceros, a *Dinotherium*, a rugose *Unio flabellatus*. The lowermost levels of this layer revealed remains of fishes, snakes, (grass-snakes?), turtles and *Limnea, Planorbis, Valvata, Bithinia* and *Pisidium* shells".

This work is the first ever mention of fossilize turtles as found not only on the territory of Georgia, but also in Azerbaijan, because the above mentioned variegated clays widely occur in the valley of the river Iori where one republic is contiguous with another. However the first description of a concrete species of extinct turtles was published only in 1960 by L.K.Gabunia and V.M.Chkhikvadze.

As on today, we can say that turtles have been relatively well studied in the following locations in Georgia.

Bazaleti (Pontian): *Testudo bosporica*.

Benara (Aquitanian): *Trionyx sp., Palaeochelys gabunii, Ergilemys meschethica.*

Dzedzvta Khevi (Meotian): *Centrochelys natadzei.*

Iori (Upper Sarmatian): *Trionyx sp., Chelydropsis sp., Mauremys sarmatica, Emydoidea tarashchuki, Testudo burtschaki.*

Kodori (Upper Pontian or Lower Kimmerian): *Sakya kolakovskii.*

Pantishara (Upper Sarmatian): *Mauremys sarmatica, Emydoidea tarashchuki, Testudo burtschaki.*

Udabno - 2 (Meotian): *Centrochelys sp., Testudo cf.eldarica, Mauremys cf.sarmatica.*

Other Tertiary and Pleistonese locations in Georgia (Metekhi, Dmanisi, Kotsakhuri, Kumros Khevi, Tori, Tsira and elsewhere) are either unstudied or represented in collections with material only too fragmentary.

Azerbaijan. We have mentioned earlier that the first reference to remnants of turtles unearthed in Azerbaijan is to be found in V.V.Bogachev's work [1,3]. The first research paper that describes and shows a fragment of the plastron of the shell of a land tortoise (unearthed in Apsheronian sediments near the village of Karasakal) has been published by N.I. Burtshcak-Abramovich [4].

The most significant research into the herpetofauna of Azerbaijan was made by A.M. Alekperov, who among other things also presents material on fossil turtles; although the author completed his studies in this line back in 1958 [5], they were published only 20 years later, in 1978 [6]. This circumstance accounts for the fact that in formal terms, it is the article by V.M. Chkhikvadze [7] that is counted as the first publication that offers description of a new species of fossil turtles in Azerbaijan.

The following locations have been studied relatively well:

Binagadi (Pleistocene): *Testudo graeca binagadenses.*

Kushkuna (Upper Akchagylian or, rather, Lower Apsheronian): *Mauremys alekperovi.*

Perekishkul (Upper Oligocene, Khattian):*Glarichelys gwinneri.*

Eldari-3 (Upper Sarmatian or, rather, Lower Meotian): *Testudo eldarica, Trionyx. sp., Mauremis sarmatica.*

Material from other locations (Boz-Dag, Duz-Dag, Enikend, Palantukan, Damdjili, Azykh, Talagar and elsewhere) has been studied insufficiently.

Armenia. In one of his reviews L.I Khozatsky [8] points out that "...remains of turtles pertaining to the genus *Clemmys*, that is close to the *Cl.caspica* have been unearthed in Pleistocene diatomites of the river Zangi (Armenia)". It is the first mention of turtles found on the territory of Armenia (at Nurnus). These materials were published only recently [9]. Besides Nurnus, only one other location is known to contain remains of *Testudo graeca armenica* - it is Verin-Khatunorkh (Holosene).

Northern Caucasus. The first mention remains of fossil turtles found in this vast region is made by A.N.Ryabinin [10] who says: "...Pleistocene deposits in the district of Pyatigorsk (travertines of Mt. Mashuk); L.V.Langwagen's collection; a rather well preserved shell of an European pond terrapin *Emys orbicularis*". The same author has also made the first description of the materials obtained from the Pre-Caucasian area [11].

Let us now consider a number of the most importaint locations:

Belomechetskaya (Middle Moicene): *Trionif. danovi, Ocadia sp.. ?Ergilemys sp., ?Protestudo sp.*

Kosyakino (Pleistocene, Moldavian faunistic complex): *Emys antiqua, Melanochelys pidoplickoi, Sakya riabinini, Testudo cernovi.*

Maikop (=Maikop I and Maikop II; Middle Sarmatian): *Trionyx khozatskyi, Emididae gen. et sp. indet.*

Nizhny Vodyanoy (Pliocene; most probably Middle or Late Pontian): *Agrionemis sp.*

Piatigorsk (Pleistosene): *Emys orbicularis.*

Chornaia Rechka (Early Miosene, most probably Burdigalian): *"Chelonia" caucasica.*

There are some other locations in the Pre-Caucasian area (Abadzekhskaya, Kutzay-Gora, Otradnaya, Taman, Tash-Kala, Elkhotovo and elsewhere), but they are still very poorly studied.

* * * * * *

Thus A.N.Ryabinin, L.I.Khozatsky, A.M.Alekperov, V.V.Bogachev, N.I.Burtschak-Abramovich, L.K.Gabunia and V.M.Chkhikvadze were pioneers of palaeochelonioIogical research in Caucasus and laid a foundation for future works in this line. Especially noteworthy is the series of publications put out by L.I. Khozatsky [8, 12-18], and V.M.Chkhikvadze [3, 7, 19-32].

In 1964 V.B.Sukhanov published his work "Subclass Testudinata" in the book "Principles of Paleontology" [33] which then was unique in the whole world regarding the scope and the depth of the material is covered. This research exerted great influence on all the subsequent investigations on the territory of the former USSR. It contains inter alia, data concerning turtles in the Caucasus.

As mentioned earlier, A.M.Alekperov's monograph "Amphibians and Reptiles of Azerbaijan" carries rather copious information about fossil turtles in the Caucasus allowing us to regard it as the first report on this subject in the above region. The above mentioned works together with the number of subsequent works were later summarized by V.M.Chkhikvadze in this monograph "Fossil Turtles of the Caucasus and Northen Black Sea Littoral" [24] and a separate article "Fossil Tortoises of the Genus *Testudo* in the USSR" [30]; see also [25, 31].

Currently three spieses of turtles live on the territory of the Caucasus: *Emys orbicularis, Mauremis caspica* and *Tescudo graeca;* the latter represented by four subspecies: *T.g.ibera; T.g.nicolskii; T.g.pallasi; T.g.armeniaka.* The variety within the species of turtles in the Caucasus in Pliocene was incomparably greater, and miocene faunas were more rich. According to the latest data [28] remains of six families of turtles were discovered on the territory of the Caucasus.

Family *Trionychidae*

Trionyx danovi Ckhikvadze, 1988.
Trionix khosatzkyi Ckhikvadze, 1983

Family *Chelydridae*

Chelydropsis sp.

Family *Cheloniidae*

Glarichelys gwinneri (Wegner, 1911).
"Chelonia" caucasia Riabinin, 1929

Family *Emididae*

Emydoidea tarashchuki Ckhikvadze, 1980.
Emys antiqua Khozatsky, 1956.

Family *Bataguridae*

Mauremys alekperovi Ckhikvadze, 1989.
Mauremys caspia gambariani Ckhikvadze, 1988.
Mauremys sarmatica (Purschke, 1885).
Melanochelys pidoplickoi (Khosatzky, 1964).
Ocadia sp. (Northern Caucasus, Middle Miosene (26).
Palacochelys gabunii Ckhikvadze, 1973.
Sakya kolakovskii Ckhikvadze, 1968.
Sakya riabinini (Khosatzky, 1964).

Family *Testudinidae*

Agrionemys sp. (Northern Caucasus, Pliocene (26)).
Centrochelys natadzei Chkhikvadze, 1989.
Ergilemys meschethica (Gabunia et Ckhikvadze, 1960).
Testudo bosporica Riabinin, 1945.
Testudo burtschaki Ckhikvadze, 1975.
Testudo cernovi Khosatzky, 1948.
Testudo cernovi transcaucasica Ckhikvadze, 1979.
Testudo eldarica Khosatzky et Alekperov, 1978.
Testudo graeca binagadenses Khosatzky, 1978.

The stratigraphic distribution of the above listed species is given in Table 1 (only Late Neogene).

Table 1

Major Localities and Characteristic Species of Turtles from Late Cenozoic Period in the Caucasus

Age	Localities	Characteristics species of turtles
Apsheronian	Boz-Dag, Duz-Dag, Enikend, Kushkuna-2, Kotsakhuri, Dmanisi.	*Emys orbicularis, Mauremys alekperovi, Testudo cf. graeca.*
Akchagylian	Kvabebi, Kumros Khevi.	*Testudo cernovi transcaucasica.*
Kimmerian	Nurnus; Kosyakino.	*Mauremys caspica gambariani, Sakya riabinini, Melanochelys pidoplickoi, Emys antiqua, Testudo cernovi*
Pontian	Kodori, Bazaleti.	*Sakya kolakovskii, Testudo bosporica.*
Meotian	Dzedzvta Khevi, Iori, Pantishara	*Centrochelys natadzei, Emydoidea tarashuchuki, Trionyx sp., Mauremys sarmatica, Testudo burtschaki, Chelydropsis sp.*
	Udabno-2. Eldari-3,	*Testudo eldarica, Testudo cf. eldarica*
Upper Sarmatian	Udabno-1. Eldari-2,	Turtles were not founded
Middle Sarmatian	Maikop-1, Maikop-2.	*Trionix khozatskyi, Emydidae indet.*

Besides the extinct forms mentioned above, remains of all the three present-day species of turtles have been found in Pleistocene and Holocene deposits in the Caucasus (these deposits are listed in greater detail in works [24,30]).

In summary, we may say the separate republic and areas in the Caucasus remain unstudied, which is eloquently attested by a small number of well studied deposits and the scantly special publications. Also noteworthy in considerable disproportion in the degree to which separate groups of fossil amphibians and reptiles have been studied. Today our information about fossil tortoises is pretty ample, while amphibians [34], and squammata reptiles [36] remain almost today uninvestigated. The very paucity of information about Mesozoic herpetofauna corroborates it too [25, 31].

Further investigation into the fossil herpetofauna of the Caucasus will allow the research to use the results obtained in it course not only to trace the morphology and philogenesis to separate groups and to determinate the time and the direction of migration, but will also offer a realistic basis for biostratigraphic correlations. Besides amphibians and reptiles are good indicators of the ecological situation, and their stenobiontic will allow us to determine reliably and precisely many parameters of the external environment of the geologic past.

The authors are greatly indebted to the George Soros Fund whose generous financial support has made the present research possible.

L.Davitashvili Institute of Paleobiology
Georgian Academy of Sciences

REFERENCES

1. *V.V. Bogachev.* Transac. Azerbaij. Branch Acad. Sci. USSR **9**, 39, 1938, 29-46 (Russian).
2. *L.K. Gabunia, V.M. Chkhikvadze.* Bull. Acad. Sci. Georg. SSR **24**, 2, 1960, 189-196 (Russian).
3. *V.V. Bogachev.* Proc. Azerbaij. Branch Acad. Sci. USSR, **6**, 1939, 91-99.
4. *N.I. Butschak-Abramovich.* Transac. Zardabi Museum Nat. History, 8, 1954, 14-49.
5. *A.M. Alekperov.* The author's Thesis of a Doctor's Dissertation, 1958, 1-26.
6. *A.M. Alekperov.* Amphibians and reptiles of Azerbaijan. Baku, 1978, 1-264.
7. *V.M. Chkhikvadze.* In: The Fauna and it's Protection in the Republics of the Transcaucasia. Erevan, 1975, 165-166.
8. *L.I. Khozatsky.* Transac. All-Russian Geographical Congress. **3**, 1949, 221-230.
9. *G.U. Melik-Adamian, E.N. Kurochkin, V.M.Chkhikvadze.* Vestnik Zoologii, **6**, 1988, 71-74.
10. *A.N.Ryabinin.* Transac. Museum Geogr. and Mineral. Acad. Sci., **1**, 1, 1918, 1-16.
11. *A.N.Ryabinin.* Transac. Russ. Mineral. Soc. **2**, 58, 2, 1929, 193-199.
12. *L.I.Khozatsky.* Voprosi Paleontologii, 1, 1951, 20-31.
13. *L.I.Khozatsky.* Ezhegodnik VPO, 15, 1956, 321-327.
14. *L.I.Khozatsky.* In: The Problems of Land Zoogeography. Lvov, 1958, 319-324.
15. *L.I.Khozatsky.* In: Stratigraphy and Paleontology of Mesozoik and Paleogene-Neogenic Continental Sediments of Asian parts of USSR, Leningrad,1967, 215-218.
16. *L.I.Khozatsky.* In: Stratigraphy of the USSR, Paleogene System. Moscow, Nedra, 1975, 439-441.
17. *L.I.Khozatsky.* In: Stratigraphi of the USSR; Quaternary System. Moscow, 1982, 248-262.
18. *L.I.Khozatsky, A.M.Alekperov.* Sci. Transac. of Azerbaijan State University, 12, 1957, 101-112.
19. *V.M.Chkhikvadze.* The 26th Students Scientific Conference. Tbilisi, 1964,15-16.
20. *V.M.Chkhikvadze.* Paleontol. J., 2, 1968, 88-94.
21. *V.M.Chkhikvadze.* Bull. Acad. Sci. Georg. SSR, **65**, 3, 1972, 745-748.
22. *V.M.Chkhikvadze.* Bull. Acad. Sci. Georg. SSR, **69**, 3, 1973, 745-749.
23. *V.M.Chkhikvadze.* Voprosy Gerpetologii, 4, 1977, 226-227.
24. *V.M.Chkhikvadze.* The Fossil Turtles of Caucasus and Black Sea Littoral. Tbilisi, 1983.
25. *V.M.Chkhikvadze.* In: Problems of Regional Fauna and Animals Ecology, Stavropol, 1987, 84-90.
26. *V.M.Chkhikvadze.* In: The Animals Kingdom of Pre-Caucasus Area and Contiguous Territories. Stavropol, 1988, 128-139.
27. *V.M.Chkhikvadze.* Neogenic Turtles of the USSR. Tbilisi, 1989,1-102.
28. *V.M.Chkhikvadze.* Paleogenic Turtles of the USSR. Tbilisi, 1990, 1-95.
29. *S.M.Aslanova, V.G.Pronin, V.M.Chkikvadze.* Vestnic Zoologii, 6, 1979, 17-22.
30. *M.A.Bakradze, V.M.Chkhikvadze.* Bull. of Georg. State Museum, 33-A, 1984, 163-174.
31. *M.A.Bakradze, V.M.Chkhikvadze.* Bull. Georg. State Museum, 34-A, 1987, 176-193.

32. *V.M.Chkhikvadze, A.N.Lungu.* In: Paleobiogeographical Researches of the Country between Rivers Dniester and Prut in Mesozonic and Cenozonic Periods. Kishinev, 1984, 72-86.
33. *V.B.Sukhanov.* In: Principles of Paleontology: Amphibians, Reptiles and Birds. Moscow, 1964, 354-438.
34. *V.M.Chkhikvadze.* Proc. Acad. Sci. Georg. SSR, ser. biol, **10**, 1, 1984, 5-13.
35. *M.B.Ephimov, V.M.Chkhikvadze.* Proc. Acad. Sci. Georg. SSR, ser. biol., **13**, 3, 1987, 200-207.
36. *G.A.Zerova, V.M.Chkhikvadze.* Proc. Acad. Sci. Georg. SSR, ser. biol., **10**, 5, 1984, 319-326.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

LINGUISTICS

I. Biniashvili

To the Meaning of the Word-Combination *Sartqeli Šapirisaj*

Presented by Academician A.Kalandadze, December 10, 1996

ABSTRACT. **The paper investigates a word-combination from the "Book of Ezekiel" which wasn't translated from Greek into Georgian in a right way. This erratum in the Greek text itself appears due to the wrong understanding of two Hebrew words being graphically identical but quite different in meanings.**

In the "Book of Ezekiel" we read *da erti ḳaci šoris matsa, šemosili qamlita da sartqeli šapirisai çelta mista* (9,2). It is under the editorship of ⴁ. At the same time under the editorship of ⴂ the corresponding place is rendered as *sartqeli sappirisaj* where *šapiri* is similar to sappiri. The word *šapiri* entered neither to Sulkhan-Saba Orbeliani nor to N. and D.Chubinashvili's Dictionaries. I.Abuladze interprets this word as *šapir-i* meaning sapphire [1]. It is based on the above mentioned chapter from the "Book of Ezekiel".

T.Tskitishvili [2] also explains the above mentioned word as *šapir-i* 9,2, whereas **sapiron-i** *η' σα'πφενροζ* means "precious stone" [2].

Compare **sapiron-i** with sapphire "honest stone"

T.Tskitishvili pays special attention to the word *šapir*. He points out that under ⴁ reduction the word *sapir* is determined by *-š-* form *sartqeli šapirisaj* (9,2, *sartqeli sappirisaj* ⴂ). It is true that the word *sapiri* is of Easten origin, but in Georgian literary tradition it is known as "sapir" (coming from Greek) [2].

Among the words of Eastern origin entered from Greek denoting particularly precious stones the author also mentions the word **sapiron** [2].

Similar word-combinations such as in Georgian *sartqeli šapirisaj* under ⴁ editorship or *sartqeli sappirisaj* under ⴂ editorship there are similar word-combinations in other languages (except Greek) meaning "the inkpot of a writer". Compare Hebrew *qeseṯ hassōpēr* , Armenian *gragiri kajamar*, Russian *chernilnitsa pistsa*. Thus here we deal with the writing equipment and not with the sapphire belt. Unlike the other languages, in Georgian we have quite another picture. In this case Georgian follows the Greek text. It is in Greek that the word combination denoting "sapphire belt" is confirmed, i.e. *ζωνη σαπφειρου.*

We suppose that the original meaning of Hebrew *qeseṭ hassoper* as sapphire belt in Greek comes from two Hebrew words having quite different meaning but at the same time approximately the similar spelling. These words are: *sōpēr* "the writer" and *sāppīr* "sapphire". The root **spr** denoting "writing" is considered to be Aramean. It entered many languages; such as Syrian **sapra**, Akadian **šâpiri** (compare Hebrew surname **šapiro**), Hebrew **sōpēr** "the writer". Compare with Hebrew *sēper* "the book, the letter" [3].

As M.Nedospasova [4] notes the word denoting *sāppir* is widely used in Hebrew Bible with the meaning of precious stone. The scientist compares it with Judish-Aramean *sampīrinow*, Syrian *sapîrā*, *satîlā* and points out that this word is considered to be borrowed from Sanscrit: *ganipгya* meaning Saturn's favourite. [4]. M.Pasmer states [5] that old Hebrew **sappîr** comes from old Indian. From Hebrew it went to Greek (>sappeiros) and then this word appeared in many other languages [4].

In Armenian the -š- form of the word sapir i.e. *šapig'* is fixed by H.Acharyan considered it to be borrowed from Assirian *sapīlā* [6].

It is evident that the source of error in Greek must be the same spelling of two Hebrew words סֹפֵר "writer" and סַפִּיר "sapphire".

This inexactness comes into Georgian from Greek. Thus we can say that the above mentioned context generally bears Greek trace. It is testified by the chapter that follows the one regarded above. In Hebrew the same word-combination *qeset hassōpēr* "the writer's inkpot" is repeated. The same is observed in Armenian and Russian. But in Greek we have only one meaning i.e. "belt". Georgian repeats Greek meaning of *romelsa igi ertqa 'sartqeli çelta mista.*

As it is seen the Georgian text of the above mentioned chapter depends on Greek.

A.Chikobava Institute of Linguistics
Georgian Academy of Sciences

REFERENCES

1. *I.Abuladze.* Old Georgian Dictionary. Tb. 1973.
2. Old Georgian Versions of the Book of Ezekiel under T.Tskitishvili editorship. Tb., 1976.
3. *W.Gesenius.* Hebräisches und Aramäisches Handwörterbuch uber das Alter Testament, Leipzig, 1912.
4. *M.E.Nedospasova.* Dragotsennie kamni v arabskoi Biblii, Tb., 1991.
5. *M.Fasmer.* Etymological Dictionary of the Russian Language III, M., 1987.
6. *H.Acharjan.* Etymological Dictionary of the Armenian Language III, Erevan, 1977.

საქართველოს მეცნიერებათა აკადემიის მოამბე, 154, №3, 1996
BULLETIN OF THE GEORGIAN ACADEMY OF SCIENCES, 154, №3, 1996

HISTORY

I.Menteshashvili

East India Company and the Great Mughals

Presented by Corr. Member of the Academy L.Chilashvili, November 11, 1996

ABSTRACT. **This article investigates the attitude of the British East India Company towards the Indians and the Great Mughals in the course of the 17th-19th centuries. Reviewing Company's business affairs, diplomatic relations and warfare we stress the change of English from self-conscious merchants to the haughty rulers and thereby the change of attitude towards natives from the respect to negligence.**

The first contacts between English and natives were far from the sense of superiority or racial predjudices. Europians were more surprised by the religion of Hindoos, then by the colour of their skin [1]. Europe of 16th-17th centuries was scanty region compared with the East. Therefore English had nothing to brag about. The wealth and pomp of Mughal court dazzled them so much that they held their clothes and their country inferior in comparison with this luxury [2]. Each gesture of attention of Mughal Emperors or courtiers was taken by Company's servants as blessing [3]. Company's Embassy to the Mughal court tried their best to carry out each requirement of native authorities [4]. Situation changed thoroughly since the battle of Plassy in 1757. If before the battle Clive is still delicate towards the Nawab of Bengal and addresses him as a great warrior and father [5] calls himself his son, after the victory of English he gives the most negative appreciation of the natives, Navab and his court. This information is given in his letter to Lawrence Sulivan, the member of the Board of Directors [6].

The attitude of English towards natives was changing along with strengthening the Company's positions in India. This is proved on a base of diaries of Company's servants since the 70s of the 18th century. Non-acceptance of vernacular values made them sure to hold British presence in India blessing for natives [7]. Being taken natives as inferiors, English dismissed them from the Company service, as it took place while Lord Cornwalliss, or Lord Wellesley's Governorship [6].

Since the Company has got an upper hand over the native Potentates and turned into actual sovereign of India English has changed their attitude towards Mughals. While Lord Wellesley's presence at office Company's resident at Mughal court still was acting in compliance with etiquette, but since 1806 they changed their attitude and gradually ceased the expression of respect to the Emperors [9].

This change was caused not only by achieving military and political superiority. The Company Raj tightened contacts of English with India and shed lustre on differences between the values of the two nations. Western pscyche, active and purposful, based on the ideas of the Renaissance [10] and Puritanism [1] did not accept passive, irrational spirit of India [12]. Achievements of Britain in the fields of economics and politics during the Industrial Revolution created notions about advanced and backward nations. All this caused negligence of English towards natives and Mughals.

Georgian State Museum

REFERENCES

1. *C.Carrington.* The British Overseas. Exploit of the Nation of Shopkeepers, 149, 1950, 154.
2. *R.Mukherjee.* The Rise and Fall of the East India Company. N.Y. 141, 1974.
3. The English Factories in India. 1624-1629. A Calendar of Documents in the India Office, Westminster. By William Forster. Oxford, 1909, 229.
4. *J.Mill.* The History of British India, **3**. L., 29-30, 1820.
5. Oriental and India Office Collections. H/193 Miscellanious., H/139.
6. *B.Gardner.* The East India Company. N.Y. 93, 140, 1990.
7. *G.Forster.* Journey from Bengal to England, through the Northern Part of India, Kashmire, Afganistan and Persia into Russia by the Caspian Sea, **1**. L., 1808., XVI.
8. *W.Ward.* A View of the History, Literature and Religion of the Hindoos, **3**. L., 1817, 242.
9. *L.M.Bhatia.* Miscellany on the Raj., 10.
10. *M.A.Barg.* Ёpokhi i Idei. Stanovlenie Istorizma. M., 1987, 237 (Russian).
11. *M.Veber.* Izbrannie Proizvedenija. M., 1990, 74 (Russian).
12. Otkritie Indii. Filosofskie i Ёsteticheskie Vozzrenija v Indii v XX v. M, 1987, 351-352 (Russian).

To Support our Journal Subscribe to it Now

Copy this form to mail or fax your order number today

Title	Volume	ISSN	Issues	Rate in US Dollars	Your order
BULLETIN of the Georgian Academy of Sciences	155 & 156	76181	6	400	$_____

1. Name______________________

Institution__________________

____

Address______________________

____

City_________________________

State________________________

Zip/Postal Code_______________

Country______________________

Telephone (_____)____________

Fax (_____)__________________

E-mail ______________________

2. My payment for $__________
is enclosed

Or

Bill me. My library's purchase order number is____________

3.Send this order and payment to:

Mrs. Ludmila Gverdtsiteli,
Editorial Manager

52, Rustaveli avenue, 380008,
Tbilisi, Georgia

Tel.: (+995-32) 99-75-93

Fax: (+995-32) 99-88-23

E-mail: Bulletin@presid.acnet.ge

The Journal will be published in 1997

გადაეცა წარმოებას 20.10.1996. ხელმოწერილია დასაბეჭდად 13.01.1997.
ფორმატი 70×108 $^1/_{16}$. აწყობილია კომპიუტერზე. ოფსეტური ბეჭდვა.
პირობითი ნაბ. თ. 10. სააღრიცხვო-საგამომცემლო თაბახი 10.
ტირაჟი 500. შეკვ. № 11 ფასი სახელშეკრულებო.

რედაქციის მისამართი: 380008, თბილისი-8, რუსთაველის პრ. 52, ტელ. 99-75-93.
საქართველოს მეცნიერებათა აკადემიის საწარმოო-საგამომცემლო გაერთიანება „მეცნიერება",
380060, თბილისი, დ. გამრეკელის ქ. 19, ტელ. 37-22-97.

INDEX 76181