„MIRCEA CEL BATRAN" NAVAL ACADEMY
SCIENTIFIC BULLETIN

Volume XI- 2008

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BULETIN ŞTIINȚIFIC

ISSN 1454 - 864X

Published by
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VOLUMUL XI – 2008

Adresa redacției:
Str. Fulgerului nr.1, 8700 Constanța
Tel. 0241/626200/1149; fax: 0241/643096
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THE ASPECTS ON THE ELECTROCONDUCTOR FLUID MOTION IN THE TOROIDAL MERCURY INDUIT GYRO-MOTOR

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PhD. Eng. V. BOZIANU

Abstract: *In this paper the authors have tried to do their modest part to the study of the electro-kinetic and magneto-hydrodynamic phenomena occurring in the operation of the liquid armature gyro-motor both the theoretical and experimental points of view.*
From the studies done by the authors it results that this mercury induit rotates in concentric layers at different speeds representing a very interesting and unknown phenomenon till now.
In the liquid armature electric machines, in this case, the mercury induit gyro-motor, an interaction between the magnetic field and an electro-conductor liquid (mercury) travelled by a current takes place. As a result of this interaction, the electric energy is converted into a mechanical energy of the fluid, mercury, which is the carrier of angular momentum, too.
Theoretically, it has been proved that this motion is special under the electromagnetic field conditions differing from the fluid motion in the pipes in the absence of electromagnetic field; this characteristic of motion was confirmed by the experimental results.
The tangential speed of the electroconductor fluid in the mercury duct is influenced by the variation of pole pitch with channel radius due to the constructional form of the toroidal stator.
Experimentally, it has been found that the mercury moves in layers, their thickness depends on the relation between the intermolecular attraction forces and Lorentz forces which put in motion the electroconductor fluid.

Keywords**:** *fluid motion, gyro-motor, toroidal mercury induit*

1 Considerations on the electroconductor fluid motion

The problem of rotating motion of electroconductor fluids near the side walls of the channels is approached in many literature works. By admitting the existence of a boundary layer of turbulent type and that the entire viscous dissipation is caused only near the walls, we can determine the conditions which establish the speed distribution in the whole operating range of mercury gyro-motors.

As compared to the classical fluid mechanics the study of electro-conductor fluid motion stability is more difficult because we must also take into account the disturbances caused by the secondary effects of the electromagnetic field.

We can maintain that the presence of this field has a stabilizing effect on the motion, this is proved by experiments

So, Murgatroyd, analyzing the experimental results obtained by Hartmann, Lazarus, Lock etc. proposed the following criterion regarding the conversion from the laminar (viscous) flow to the turbulent flow, inside the boundary layer:

$$R_{e_{cr}} = 250\frac{H_a^2 \cdot th(H_a)}{H_a - th(H_a)} \tag{1}$$

where:

H_a = is Hartmann's number given by the expression:

$$H_a = B \cdot R_h \sqrt{\frac{\sigma}{2\eta}}$$

In which:

B = magnetic induction;

R_h = equivalent hydraulic radius;

σ = fluid electro-conductibility;

η = dynamic viscosity.

Branover and Lielausis, by their researches, confirmed the justness of this proposed criterion. So, by studying the electro-conductor fluid flow in the presence of electromagnetic field and by passing the semi-empirical laws of Prandtl type by Harrisi, they obtained a calculation relation for the hydraulic loss factor:

$$\lambda = \lambda_0(1 + 2S) \tag{2}$$

where:

λ_0 = is the hydraulic loss factoring the absence of fields ($\vec{B}, \vec{E} = 0$)

S = Stuart's number

$$S = \frac{H_a^2}{R_e} \tag{3}$$

Among the patterns proposed by different authors, the nearest one to our pattern is the one which admits the existence of a friction layer of "disk-crown" type in the rotational motion at which the fluid has constant properties, the field is homogeneous and parallel to the disk axis.

We admit the existence of only one speed component, the tangential speed which varies with the radius according to a law given by the expression:

$$v_\theta = r^n \tag{4}$$

The flow is considered to be stationary and symmetrical about an axis, so:

$$\frac{\partial v}{\partial t} = 0 \ ; \quad \frac{\partial v}{\partial \theta} = 0$$

Considering the cylindrical coordinate system (r, θ, x) with the axis x passing through the disk centre and parallel with induction $\vec{B}$, the fluid having its density ρ, its kinetic viscosity υ and its electric conductibility σ constant, we can introduce the dimensionless values:

$$\xi = \frac{\sigma \cdot B_0^2 \cdot r}{\rho \cdot v_\theta} \ ; \ \mu = \frac{x}{\delta_M} \ ; \ \sigma_M = \sqrt{\frac{\rho \cdot \upsilon}{\sigma \cdot B_0^2}} \ ; \ F_r = \frac{v_r}{v_0} \ ; \ F_\theta = \frac{v_\theta}{v_0} \ ;$$

$$F_x = \frac{\delta_M \cdot v_x}{\upsilon}; \quad g_1 = \frac{j_r}{\sigma v_0 B_0}; \quad g_2 = \frac{j_\theta}{\sigma v_0 B_0}; \quad g_3 = \frac{\delta_M \cdot j_x}{\upsilon \sigma B_0}; \tag{5}$$

where:

-υ_r, υ_0, υ_x are the speed components in the boundary layer;

-j_r, j_0, j_x are the current densities related to the three axes.

Now, we can write the dimensionless equations of the boundary layer, such as:

$$\begin{gathered}
(1-n)F_r \frac{\partial F_r}{\partial \xi} + \frac{n}{\xi} F_r^{\,2} + F_x \frac{\partial F_r}{\partial \mu} + \frac{1}{\xi}(1-F_\theta^2) = \frac{\partial^2 F_r}{\partial \mu^2} - F_r \\
(1-n)F_r \frac{\partial F_\theta}{\partial \xi} + \frac{1+n}{\xi} F_r F_\theta + F_x \frac{\partial F_\theta}{\partial \mu} = \frac{\partial^2 F_\theta}{\partial \mu^2} - (1-F_\theta) \\
(1-n)\frac{\partial F_r}{\partial \xi} + \frac{1+n}{\xi} F_r + \frac{\partial F_x}{\partial \mu} = 0
\end{gathered} \tag{6}$$

Putting the limiting conditions:

$$\begin{gathered}
F_r = 0;\ F_\theta = 0;\ F_x = 0;\ g_3 = 0 \ \text{ pentru } \ \mu \to 0 \\
F_r = 1;\ F_\theta = 1; \ \text{ pentru } \ \mu \to \infty
\end{gathered} \tag{7}$$

By eliminating F_x from the first three equations and by integrating from 0 to ∞ for the limiting conditions we obtain the following equations:

$$\begin{cases}
(1-n)\dfrac{\partial}{\partial \xi}\displaystyle\int_0^\infty F_r^2 d\mu + \dfrac{2n+1}{\xi}\displaystyle\int_0^\infty F_r^2 d\mu + \dfrac{1}{\xi}\displaystyle\int (1-F_\theta^2) d\mu = \left(\dfrac{\partial F_r}{\partial \mu}\right)_0 - \displaystyle\int_0^\infty F_r d\mu \\
(1-n)\dfrac{\partial}{\partial \xi}\displaystyle\int_0^\infty F_r(1-F_\theta) d\mu + \dfrac{1+n}{\xi}\displaystyle\int_0^\infty F_r(1-2F_\theta) d\mu = -\left(\dfrac{\partial F_\theta}{\partial \mu}\right) - \displaystyle\int_0^\infty (1-F_\theta) d\mu
\end{cases} \tag{8}$$

The solution of the system (8) is determined for $\xi \geq 1$ under the form of a power series with negative exponent of ξ. The first terms are kept, so we can write:

$$\begin{cases}
F_r = -\dfrac{1}{\xi}\left[\left(\mu - \dfrac{1}{3}\right)e^{-\mu} + \dfrac{1}{3}e^{-2\mu}\right] \\
F_\theta = 1 - e^{-\mu}
\end{cases} \tag{9}$$

-F_θ is a function with monotone variation with which we can determine the dimensionless value:

$$\delta = \int_0^\infty (1 - f_\theta) d\mu \tag{10}$$

that marks the motion layer. We can note that its value is 1 if we use the dimensionless variables chosen.

Considering those mentioned above, we admit as the parameters of the motion layer $\delta = \dfrac{\delta^*}{\delta_M}$ and α, defining the maximum value of the function F_r, so that the solutions (9) can be written under the following form:

$$\begin{cases} F_r = \dfrac{\alpha}{\xi}\left[\left(\dfrac{\mu}{\delta}-\dfrac{1}{3}\right)e^{\frac{\mu}{\delta}}+\dfrac{1}{3}e^{-\frac{2\mu}{\delta}}\right] \\ F_\theta = 1-e^{-\frac{\mu}{\delta}} \end{cases} \tag{11}$$

The parameters α and δ will be functions of ξ and they will have to prove the limiting conditions.

By introducing the solutions (11) into (6) we obtain the system:

$$\begin{cases} (1-n)\dfrac{\alpha\delta^2}{\xi}\cdot\dfrac{d\alpha}{d\xi}+\left(\dfrac{37}{7}n+\dfrac{16}{7}\right)\dfrac{\alpha^2\delta^2}{\xi^2}+\dfrac{\alpha\delta}{7}-\dfrac{64}{7}\alpha+9\delta^2=0 \\ \dfrac{1-n}{2}\dfrac{\alpha}{\xi}\dfrac{d}{d\xi}(\delta^2)-\left(\dfrac{47}{7}n+\dfrac{39}{7}\right)\dfrac{\alpha^2\delta^2}{\xi^2}-\dfrac{37}{7}\alpha\delta^2+\dfrac{100}{7}\alpha-9\delta^2=0 \end{cases} \tag{12}$$

For n = 1, the system (12) becomes an algebraical system with the following solutions:

$$\begin{cases} \alpha_0 = \sqrt{(8{,}1S^2)+2(8{,}1S^2)}-(8{,}1S^2) \\ \delta_0 = \dfrac{1}{1+\dfrac{8}{9}\cdot\dfrac{\alpha}{S^2}} \end{cases} \tag{13}$$

where: S is the Stuart's number given by the expression (3).

For n ≠ 1, the equations for calculating α and δ represent a non-linear differential system of first-order which is difficult to solve analytically.

Using a progressive development for α and δ depending on ξ, the following system is obtained:

$$\begin{cases} \alpha = 1+\dfrac{A_1}{\xi}+\dfrac{A_2}{\xi}+\dfrac{A_3}{\xi}+\ldots+\dfrac{A_i}{\xi^i} \\ \delta = 1+\dfrac{B_1}{\xi}+\dfrac{B_2}{\xi}+\dfrac{B_3}{\xi}+\ldots+\dfrac{B_i}{\xi^i} \end{cases} \tag{14}$$

By substituting the expression (14) in the system (13), an algebraical system is obtained for the calculations of the factors A_i ,B_i with the following solutions:

$$\begin{cases} A_1 = B_1 = 0 \\ A_2 = -\dfrac{32}{81}+\dfrac{n}{3}; \\ B_2 = -\dfrac{23}{36}-\dfrac{n}{4}; \\ A_3 = 0; \\ B_3 = 0; \\ A_4 = \dfrac{5n-37}{36}A_2-\dfrac{2n+30}{36}B_2 \end{cases} \tag{15}$$

For low values of ξ, the system can be numerically integrated. The results of integration for the values of n = 1; 0.5; 0; -0.5; -1 are presented in the diagrams in the Figures 1 and 2.

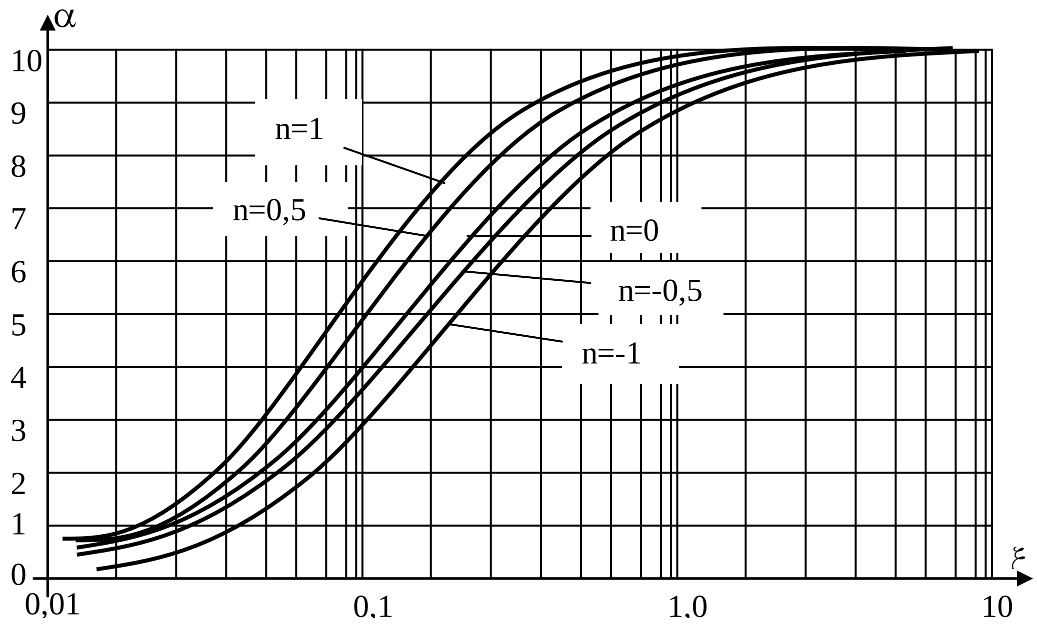

Figure 1 Results of integration α = f(ξ)

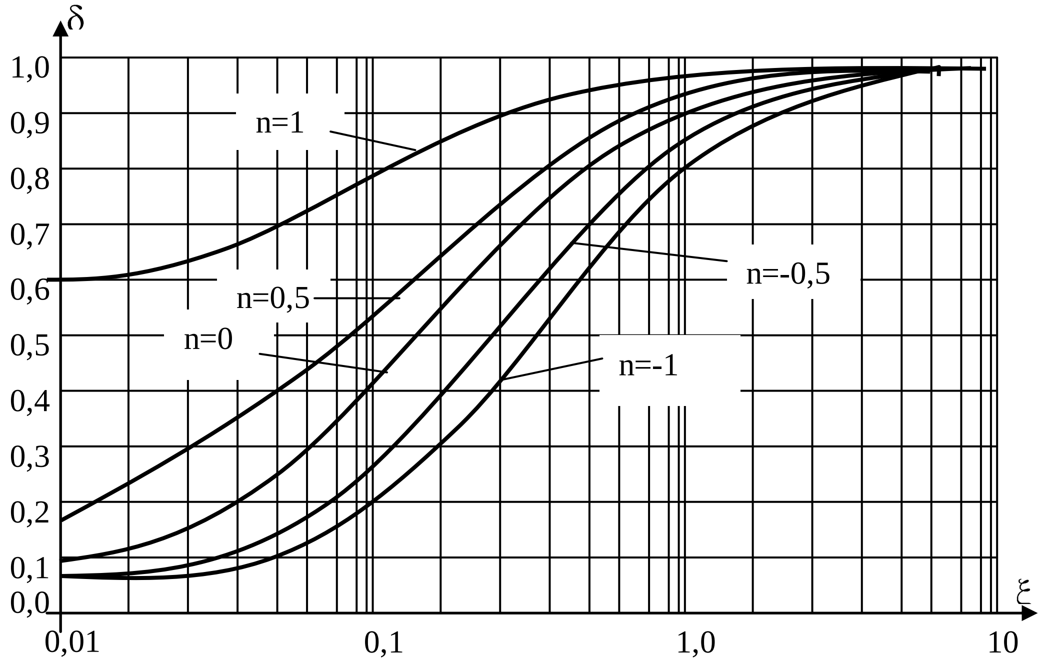

Figure 2 Results of integration δ = f(ξ)

The expression of tangential speed for a gyromotor with a double toroidal inductor and a mercury disk-armature has the form:

$$v_{\theta} = \omega_0 (r + A \cdot r^n + B \cdot r^{-n}) \tag{16}$$

with its diagram shown in Figure 3.

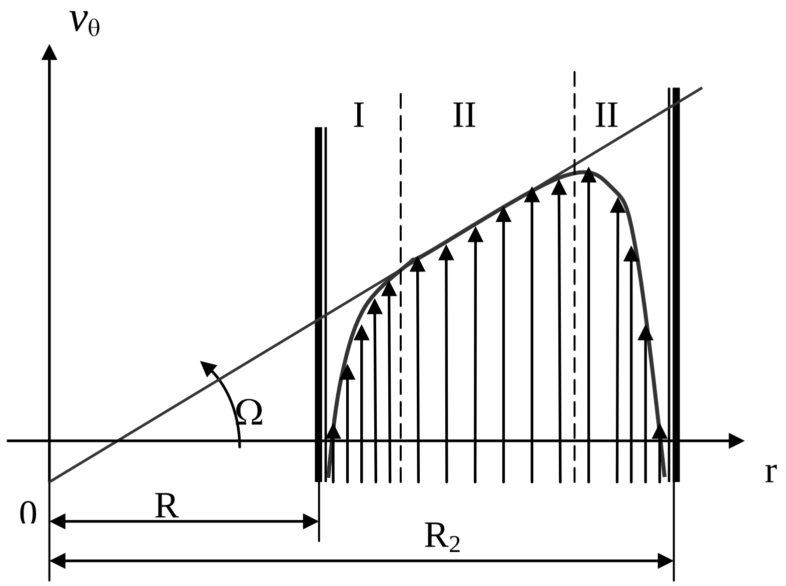

Figure 3 – The theoretical profile of fluid speeds related to the duct rings

2 The generalities on the speed distribution measurement

In the special literature of classical hydrodynamics there are presented different methods and means for measuring the fluid speed. In order to analyze the possibility of adaptation of such device for measuring the speed of mercury in the gyromotor duct, these devices will be briefly presented, which, according to their operating principle, can be:

a) the devices with an operation based on the measurement of pressure difference, they are built so that they should determine a pressure loss which can be measured and correlated with the fluid speed. The pressure drop can be determined by the changes of kinetic energy, by surface friction, due to the form friction or to their combinations.

b) the mechanical devices which have as a main element, a movable part (float, disk, turbine) moving or rotating at a speed determined by the speed of the fluid in which that element is immersed.

c) the thermal devices with an operation based on the correlation between the heat loss of a warm wire or the temperature rise of a fluid at its passing through a filament resistance (rheostat) and the fluid speed.

d) the electromagnetic devices with an operation based on the law of electromagnetic induction.

e) the ultrasonic devices with an operation based on the measurement of time difference (delay) appearing at the propagation of some acoustical pulsations on a certain distance to the same direction opposed to the motion of fluid through the pipe.

f) the radioactive devices with an operation based on the generation of radioisotopes by a tablet of radioactive substance placed in a fluid of which speed must be measured.

From the analysis of these methods we have been found that for mercury they are improper.

Taking into account the shape and small dimensions of the mercury duct (Fig.4), we have to use a small instrument for a minimum disturbance of flow.

Further on, we shall present the most proper device used for measuring the mercury speed in the gyromotor.

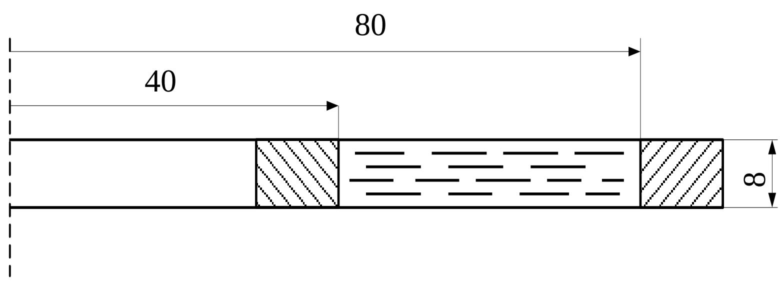

Figure 4 – The dimensions of mercury duct

3. The Pitot tube device

The operating principle of the Pitot tube is shown in Figure 5. This device is composed of two concentric tubes disposed in a parallel direction with the flow, connected to a differential pressure gauge. The outside tube is provided with small holes on its lateral surface, normally to the direction of flow and connected to the annular space.

The fluid pressure in these holes is the static pressure, ρ_1 , and so, the annular space between the two tubes is useful for transmitting the static pressure.

The inner tube is provided with a small hole at the frontal end which is useful for transmitting the total pressure (static and dynamic pressures). In the inner tube (2) the fluid raises to a level corresponding to the total load, H_1.

$$H_1 = \frac{P_s}{\gamma} + \frac{v^2}{2g} \tag{17}$$

The liquid head, H_1, is kept in equilibrium by the total pressure, P_t, which results from the static pressure, P_s, adding the pressure resulted from the kinetic energy conversion $V^2/2g$ into potential energy.

$$P_t = P_s + \rho\frac{v^2}{2} \tag{18}$$

The term $\rho\frac{v^2}{2}$ is named the impact or dynamic pressure.

In the outside tube (1) the liquid raises to a pressure head corresponding to the static pressure.

$$H_2 = \frac{P_s}{\gamma} \tag{19}$$

The level difference, ΔH, is given by the relation:

$$\Delta H = H_1 - H_2 = \frac{v^2}{2g} \tag{20}$$

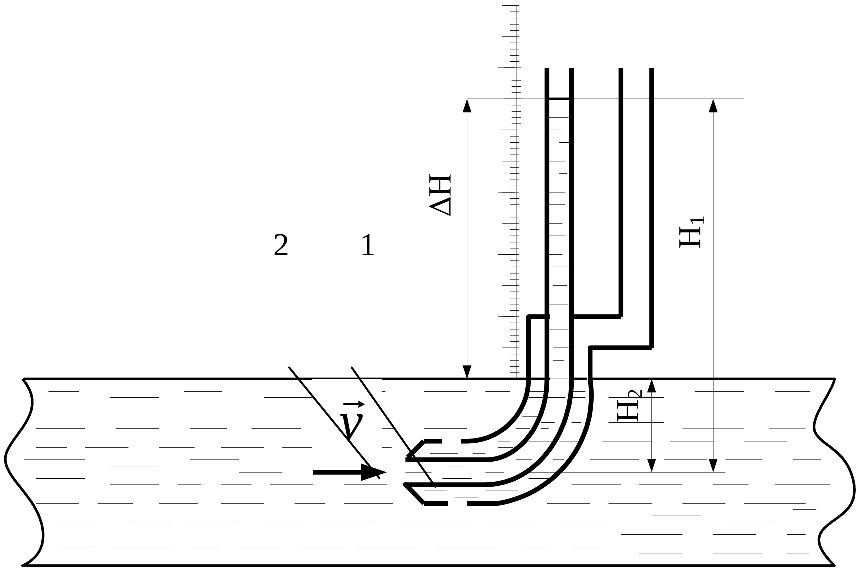

Figure 5 – The Pitot tube device

The fluid speed is given by the relation:

$$v = \sqrt{2g(\Delta H)} \tag{21}$$

The equation (21) is applied to the fluid flow (when V < 70m/s); at higher speeds the value ΔH is multiplied by a correction factor which takes into account the fluid compressibility.

Mercury is not an ideal fluid and so, the relation (21) is changed as follows:

$$\Delta H = K_x \frac{V^2}{2g} \tag{22}$$

where: K_x is the parameter which depends on the mercury viscosity and on the turbulence effects appearing near the Pitot tube placed in the moving fluid and so, on the mercury speed, too.

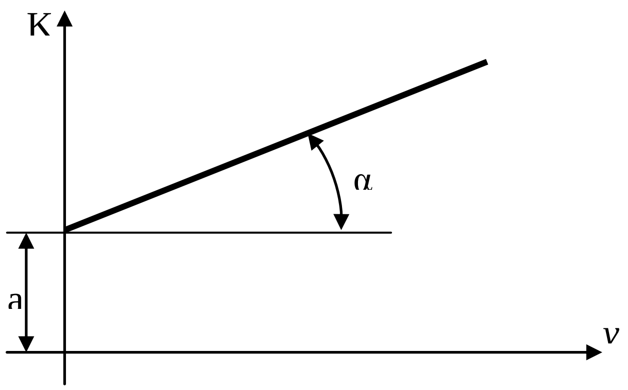

Figure 6 – The variation of parameter K_x with the fluid speed

Under these conditions, the parameter K_x is given by the expression:

$$K_x = a+bv \tag{23}$$

$$b = tg\,\alpha$$

where: a is the constant which depends on the shape of the tube.

In order to use this method, it was made a previous calibration under the similitude conditions for determining the value of motion resistance factor, K_x.

Therefore, it was built a device by means of which the variation of the parameter K_x was determined, with a diagram presented in Figure 7.

After the calibration of the Pitot tube and the determination of the parameter K_x, the measurements at the mercury gyromotor were made for establishing the rotational speed of mercury armature.

The local linear speed for a certain fluid layer was determined with the expression:

$$v = \sqrt{\frac{2g(\Delta H)}{K_x}} \tag{24}$$

and the angular speed with the expression:

$$\omega = \frac{v}{R} \tag{25}$$

where: υ = the local speed for a certain layer corresponding to the radius R

R = the radius of mercury layer.

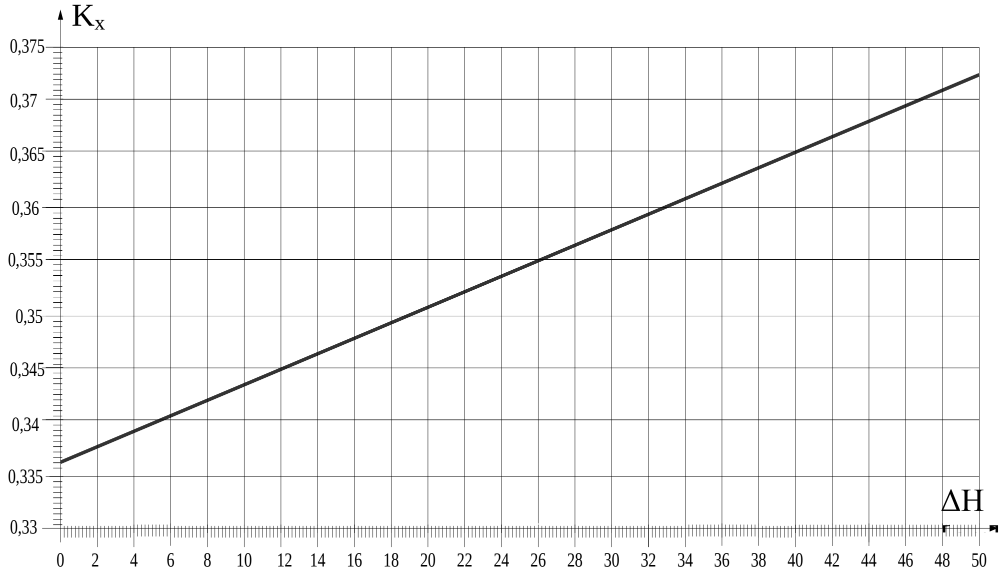

Figure 7 – The variation of parameter K_x with the mercury level/height in the Pitot tube – experimental results

4 The experimental results

The tests were made for the frequencies of 330 Hz and 500 Hz at the air gaps of 8mm and at different radii of 45, 52, 60, 67, 75 mm.

The testing results are presented in the diagrams of Figure 8- 10 where the variations of local tangential speed, v_0, are plotted at different radii at f = 330 Hz and f = 5oo Hz.

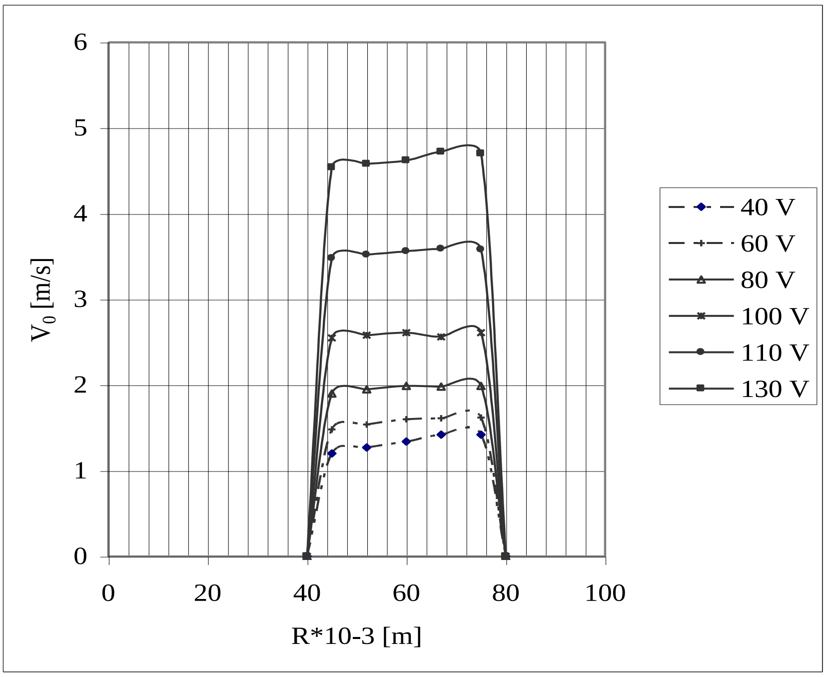

Figure 8– The variation of local tangential speed in the mercury layers with the radius, at different supply voltages at f = 330 Hz.

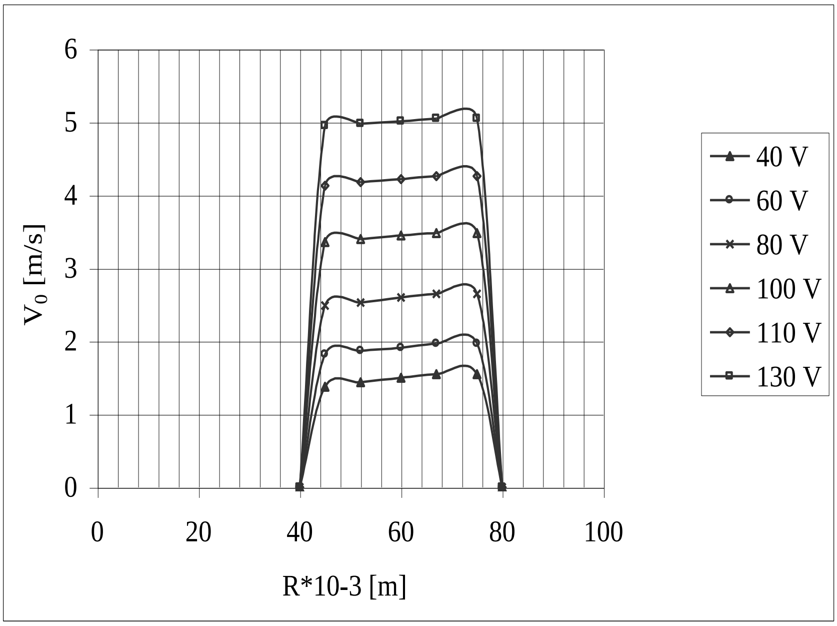

Figure 9– The variation of local tangential speed in the mercury layers with the radius, at different supply voltages at f = 500 Hz.

5. Conclusion

It has been found that the tangential speeds increase with the radius, which it is normal, and not proportional to this, but slower.

So, while the ratio of the extreme radii at which the speeds were measured is 75/45 = 1.66, the ratio of local tangential speeds is of 1.42/1.2 = 1.18.

The calculation of angular speeds represented in Figures 7 and 8, shows a very interesting phenomenon namely, the mercury rotates at different angular speeds, higher at smaller radii and lower at bigger radii.

It results that the mercury moves in layers at different speeds, but in a different way than it is known in the literature, represented in Figure 1.

It can be considered a mean rotational speed at an average radius of 60mm corresponding to a rotational speed n = 416.86 rev/min at 330Hz and to a rotational speed n = 495.65 rev/min at 500Hz.

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DYNAMIC MODELING OF THE FLOW OF AN UNDERWATER REMOTELY OPERATED VEHICLE (ROV) USED AGAINST MARITIME MINEFIELDS

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Abstract: *Maritime mine countermeasures represents, as a whole, one of the most important and sensitive actions, in either naval warfare or asymmetric warfare. Due to their relatively simple construction, maritime mines can be effective means within the terrorists' arsenal in order to block the maritime communications. Taking in account the underwater remotely operated vehicles' capabilities (ROV) they are increasingly used for maritime areas surveys as well as for mine clearance. The present paper describes the results of a ROV's motion modeling gathered by the authors while working out a research and development project carried out in the „Mircea cel Bătrân" Naval Academy of Constanta, Romania. Following the quantitative expressions of the differential equations of motion through the water, the authors successfully modeled the ROV dynamics based on Fluent software.*

Keywords: *ROV, modeling, motion, equations*

1. Introduction

ROV and the surrounding liquid that moves through form together a complex hydrodynamic system. Unlike the motion of bodies through air, the motion equations through water will rely on the considered hypothesis regarding the water around ROV. The liquid surrounding the ROV can be divided into three zones:

1. The laminar flow zone: the liquid is considered to be real and compressible. The viscous forces are present here.
2. The wake or the turbulent flow zone: beyond a certain value of the speed of the fluid particle against ROV, its movement is only possible by trajectory changing. For the streamlined bodies the actual spot where the limit stratus peels off (the point the turbulent flow starts from) is located in the vehicle's stern region (the conical segment).
3. The external potential flow zone: here, the liquid is considered to be ideal. Because the laminar stratus' thickness is pretty small, the liquid within the third zone influences the ROV's movement through water. This layer becomes thinner as the speed grows.

2. The equations of motion

When the ROV travels at a constant velocity the quantity of movement (the momentum) induced to the fluid particles has a constant value. Considering the momentum variation theorem and the angular momentum variation theorem, the energetic action of the water upon the ROV will be null.

In the case of unstable movement, the variation of momentum is permanent, which determines the appearance of the inertial forces and torques.

The main hypothesis that stands at the bottom of this phenomenon claims that the moving ROV and the volume of liquid which acts on it form altogether a complex hydrodynamic system. In other words the movement of the ROV's mass center on the desired trajectory has to consider the hydrodynamic "behavior" of both ROV and the external liquid volume it pushes.

In these conditions the theorems of momentum and angular momentum will have the following expressions:

$$\frac{dg(H+Hi)}{dt}=F \qquad \frac{dg(K+Ki)}{dt}=M \qquad (2.1)$$

where:

H, H_i = momentums of the ROV itself and of the surrounding liquid, respectively, expressed in the fixed coordinate system;

K, K_i = angular momentums of the ROV itself and of the surrounding liquid, respectively, expressed in the fixed coordinate system;

F = sum of external forces applied to the hydrodynamic system

M = sum of the external torques.

According to reference [2], the equations (2.1) can analyze ROV's motion through the ideal liquid, in the conditions of including inside the vectors F and M the supplementary forces and torques that act in this case. The motion equations (2.1) can be expressed also in the relative coordinate system, attached to the ROV's body. In such case they have a simplified form.

It means that the supplementary modification of the torque will equal (vdt) x F and the gradient of torque v x F becomes:

$$\frac{dg(H+H_1)}{dt}=\frac{d(H+H_1)}{dt}+\omega x(H+H_1)=F$$

$$\frac{dg(K+K_1)}{dt}=\frac{d(K+K_1)}{dt}+\omega x(K+K_1)+vx(H+H_1)=M \qquad (2.2)$$

where $v \times (H + H_1)$ represents the derivative of torque in time.

The equations (2.2) projected on the relative reference system will have the expressions:

$$
\begin{aligned}
&\frac{d(H_x + H_{x1})}{dt} + \omega_y(H_z + H_{z1}) - \omega_z(H_y + H_{y1}) = F_x \\
&\frac{d(H_y + H_{y1})}{dt} + \omega_z(H_x + H_{x1}) - \omega_x(H_z + H_{z1}) = F_y \\
&\frac{d(H_z + H_{z1})}{dt} + \omega_x(H_y + H_{y1}) - \omega_y(H_x + H_{x1}) = F_z \\
&\frac{d(K_x + K_{x1})}{dt} + \omega_y(K_z + K_{z1}) - \omega_z(K_y + K_{y1}) + v_y(H_z + H_{z1}) - v_z(H_y + H_{y1}) = M_x \\
&\frac{d(K_y + K_{y1})}{dt} + \omega_z(K_x + K_{x1}) - \omega_x(K_z + K_{z1}) + v_z(H_x + H_{x1}) - v_x(H_z + H_{z1}) = M_y \\
&\frac{d(K_z + K_{z1})}{dt} + \omega_x(K_y + K_{y1}) - \omega_y(H_x + H_{x1}) + v_x(H_y + H_{y1}) - v_y(H_x + H_{x1}) = M_z
\end{aligned}
\tag{2.3}
$$

Equations (2.3) represent the ROV's motion equations through the ideal liquid. Taking in account this equations the kinetic energy of ROV will be:

$$
\begin{aligned}
T_1 = {} & 1/2 \cdot \sum m_i (v_x{}^2 + v_y{}^2 + v_z{}^2) + \sum m_i z(v_x\omega_y - v_y\omega_x) + \\
& + \sum m_i y(v_z\omega_x - v_x\omega_z) + 1/2 \cdot \omega_x{}^2 \cdot \sum m_i (y^2 + z^2) + \\
& + 1/2 \cdot \omega_y{}^2 \cdot \sum m_i (x^2 + z^2) + 1/2 \cdot \omega_z{}^2 \cdot \sum m_i (x^2 + y^2) - \\
& - \omega_x\omega_y \cdot \sum m_i xy - \omega_x\omega_z \cdot \sum m_i xz - \omega_y\omega_z \cdot \sum m_i yz
\end{aligned}
\tag{2.4}
$$

The kinetic energy of the fluid may be expressed as:

$$T_2 = \rho/2 \cdot \iiint_{(w)} V^2 dw$$

where the integration considers the entire volume "w" full of liquid. Considering the velocities potential function φ, the kinetic energy is:

$$T_2 = \rho/2 \cdot \iiint_{(w)} [(\partial\varphi/\partial x)^2 + (\partial\varphi/\partial y)^2 + (\partial\varphi/\partial z)^2] dw$$

Hence, the liquid's kinetic energy has the expression:

$$
\begin{aligned}
T_2 = {} & 1/2 \cdot \lambda_{11} v_x{}^2 + 1/2 \cdot \lambda_{22} v_y{}^2 + 1/2 \cdot \lambda_{33} v_z{}^2 + \\
& + 1/2 \cdot \lambda_{26} v_y \omega_z + 1/2 \cdot \lambda_{44} \omega_x{}^2 + 1/2 \cdot \lambda_{55} \omega_y{}^2 + 1/2 \cdot \lambda_{66} \omega_z{}^2
\end{aligned}
\tag{2.5}
$$

From symmetry reasons, $\lambda_{ik} = \lambda_{ki}$ so as $\lambda_{35} = \lambda_{53}$ and $\lambda_{62} = \lambda_{26}$

The ROV's speed components in the speed coordinate system are v_x, v_y, v_z and are functions of speed v, attack angle α and drift angle β. These components will be:

$$
\begin{aligned}
v_x &= v \cdot \cos\alpha \cdot \cos\beta \\
v_y &= v \cdot \sin\alpha \cdot \sin\beta \\
v_z &= v \cdot \sin\beta
\end{aligned} \tag{2.6}
$$

Based on the previous equations, the motion equations system becomes:

$$
\begin{aligned}
& m \cdot \frac{dv_x}{dt} + (m+\lambda_{33}) \cdot \omega_y v_z - (m+\lambda_{22}) \cdot \omega_z v_y - \lambda_{26} \cdot (\omega_y{}^2 + \omega_z{}^2) = F_x \\
& (m+\lambda_{22}) \cdot \frac{dv_y}{dt} + m \cdot \omega_z v_x - (m+\lambda_{33}) \cdot \omega_x v_z + \lambda_{26} \cdot (\frac{d\omega_z}{dt} - \omega_x \omega_y) = F_y \\
& (m+\lambda_{33}) \cdot \frac{dv_y}{dt} + m \cdot \omega_y v_x + (m+\lambda_{22}) \cdot \omega_x v_y + \lambda_{26} \cdot (\frac{d\omega_z}{dt} - \omega_x \omega_z) = F_z \\
& (J_x + \lambda_{44}) \cdot \frac{d\omega_x}{dt} = M_x \\
& (J_y + \lambda_{55}) \cdot \frac{d\omega_y}{dt} + (J_x + \lambda_{44} - J_z - \lambda_{66}) \cdot \omega_x \omega_z - \lambda_{26} (\frac{dv_z}{dt} + \omega_x v_y - \omega_y v_x) = M_y \\
& (J_z + \lambda_{66}) \cdot \frac{d\omega_z}{dt} + (J_y + \lambda_{55} - J_x - \lambda_{44}) \cdot \omega_x \omega_y - \lambda_{26} (\frac{dv_y}{dt} + \omega_x v_z - \omega_z v_x) = M_z
\end{aligned} \tag{2.7}
$$

Introducing (2.5) in (2.7) we will have:

$$
\begin{aligned}
& mv \cdot \cos\alpha \cdot \cos\beta - mv \cdot \sin\alpha \cdot \cos\beta \cdot \dot{\alpha} - mv \cdot \cos\alpha \cdot \sin\beta \cdot \dot{\beta} - \\
& -(m+\lambda_{22})\omega_z v \cdot \sin\alpha \cdot \cos\beta - \lambda_{26} \cdot (\omega_y{}^2 + \omega_z{}^2) + (m+\lambda_{33}) \cdot \omega \cdot v \cdot \sin\beta = F_x \\
& -(m+\lambda_{22}) \cdot \dot{v} \cdot \cos\alpha \cdot \cos\beta - (m+\lambda_{22}) \cdot v \cdot \cos\alpha \cdot \cos\beta \cdot \dot{\alpha} + \\
& +(m+\lambda_{22}) v \cdot \sin\alpha \cdot \sin\beta \cdot \dot{\beta} + m\omega_z v \cdot \cos\alpha \cdot \cos\beta - (m+\lambda_{33}) \cdot \omega_x \cdot v \cdot \sin\beta + \\
& +\lambda_{26} \cdot (\dot{\omega}_z - \omega_x \omega_y) = F_y \\
& (m+\lambda_{33}) \cdot \dot{v} \cdot \sin\beta + (m+\lambda_{33}) \cdot v \cdot \cos\beta \cdot \dot{\beta} - (m+\lambda_{22}) \omega_x v \cdot \sin\alpha \cdot \cos\beta - \\
& -m\omega_y v \cdot \cos\alpha \cdot \cos\beta - \lambda_{26} \cdot (\dot{\omega}_y - \omega_x \omega_z) = F_z \\
& (J_x + \lambda_{44}) \cdot \dot{\omega}_x = M_x \\
& (J_y + \lambda_{55}) \cdot \dot{\omega}_y + (J_x + \lambda_{44} - J_z - \lambda_{66}) \cdot \omega_x \omega_z - \lambda_{26} (\dot{v} \cdot \sin\beta + v \cdot \cos\beta \cdot \dot{\beta} - \\
& -\omega_x v \cdot \sin\alpha \cdot \cos\beta) = M_y \\
& (J_z + \lambda_{66}) \cdot \dot{\omega}_z + (J_y + \lambda_{55} - J_x - \lambda_{44}) \cdot \omega_x \omega_y - \lambda_{26} (-\dot{v} \cdot \sin\alpha \cdot \cos\beta - \\
& v \cdot \cos\dot{\alpha} \cdot \cos\beta + v \cdot \sin\alpha \cdot \sin\beta \cdot \dot{\beta} - \omega_x v \cdot \sin\beta + \omega_z v \cdot \cos\alpha \cdot \cos\beta) = M_z
\end{aligned} \tag{2.8}
$$

3. Modeling of the flow around the ROV's body

Based on the facilities offered by the numerical integration software Fluent, the distribution of the liquid's velocities and pressures on the hull are presented below. We considered a stationary and subsonic flow. Heat exchange and gravity forces have been neglected. The surface asperities are 5μm high.

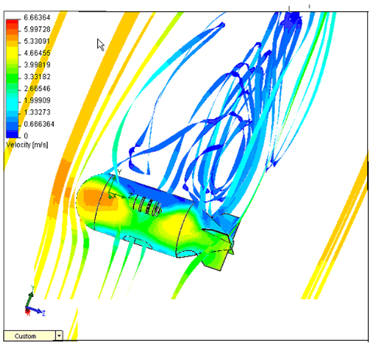

Fig.3.1 Current lines (Lateral flow)

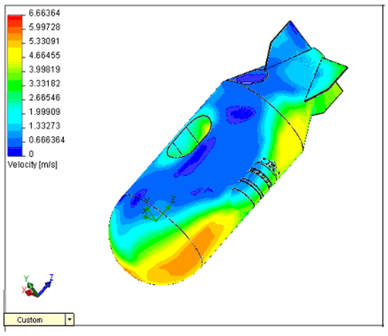

Fig.3.2 Velocities distribution (Lateral view)

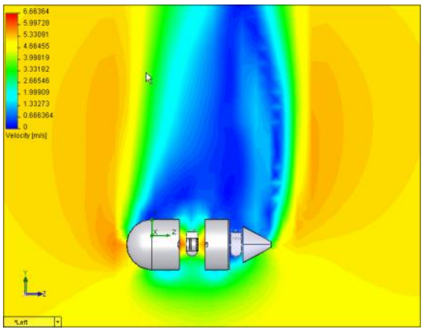

Fig. 3.3 Current lines (Frontal flow)

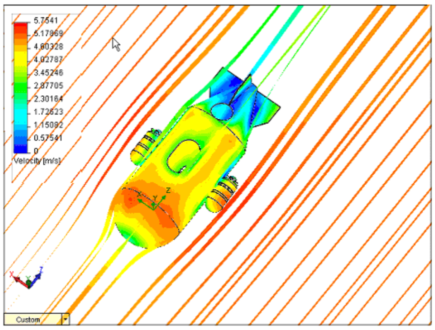

Fig. 3.4 Pressure distribution

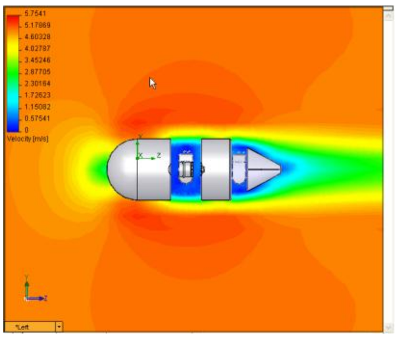

Fig. 3.5 Velocities distribution (Axial flow)

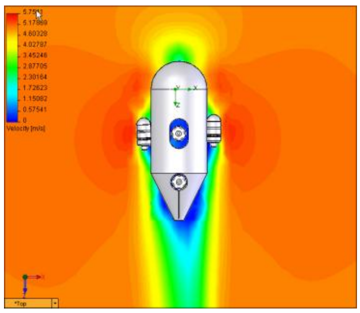

Fig. 3.6 Velocities distribution (Horizontal plane)

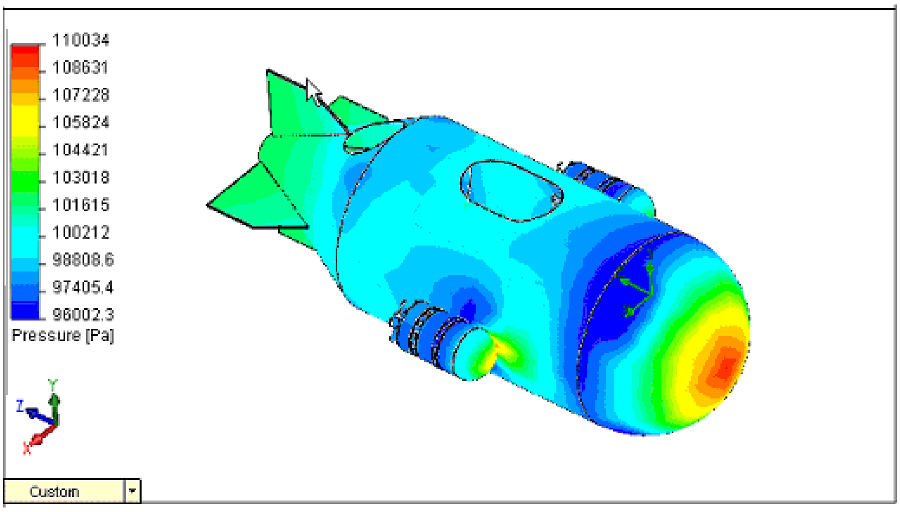

Fig. 3.7 Pressure distribution on ROV's surface

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[4] *** *User's guide* COSMOSFlow

SENSORLESS CONTROL OF PERMANENT MAGNET SYNCHRONOUS MOTOR (PMSM) BY THE AID OF A REFERENCE ADAPTIVE SYSTEM

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***Abstract**: Speed and torque controls of permanent magnet synchronous motors are usually attained by the application of position and speed sensors. However, speed and position sensors require the additional mounting space, reduce the reliability in harsh environments and increase the cost of a motor. Therefore, many studies have been performed for the elimination of speed and position sensors. This work investigates a novel speed sensorless control of a permanent magnet synchronous motor. The proposed control strategy is based on a Reference Adaptive System (RAS) using the state observer model with the current error feedback and the magnet flux model. The proposed algorithm has been verified through the simulation and experiment.*

***Keywords**: PMSM, Reference Adoptive System*

1. Introduction

The vector control in the speed and torque controlled AC drive is widely used for a high performance application. The vector control of a permanent magnet synchronous motor is usually implemented through measuring the speed and position. However, speed and position sensors require the additional mounting space, reduce the reliability in harsh environments and increase the cost of a motor. Various control algorithms have been proposed for the elimination of speed and position sensors: estimators using state equations, Luenberger or Kalman-filter observers, sliding mode control, saliency effects, artificial intelligence, direct control of torque and flux, and so on [1-4]. This paper proposes the control strategy based on the Reference Adaptive System in the sensorless control of a permanent magnet synchronous motor. This algorithm is well-known in the sensorless control of an induction motor, and has been proved to be effective and physically clear [5-7]. The proposed algorithm is verified through the simulation and experiment.

2. Mathematical Modeling of the Permanent Magnet Synchronous Motor (PMSM)

Fig. 1 shows the equivalent model of a permanent magnet synchronous motor. R_e and L_e in Fig. 1 indicate the equivalent resistance and inductance. Flux reference axes are also shown in Fig. 1.

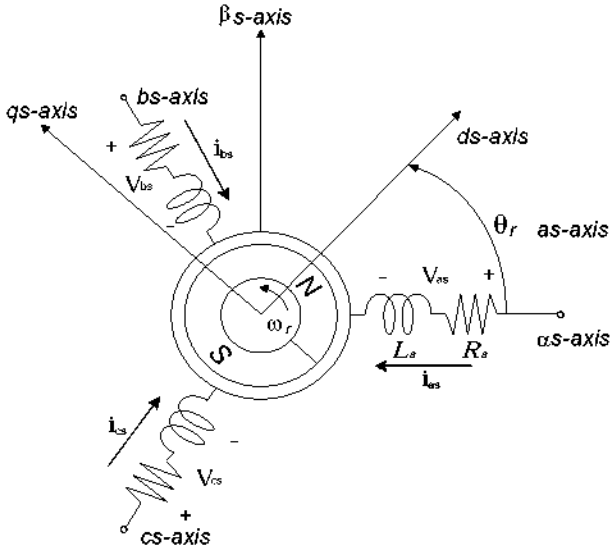


Fig. 1 The Equivalent model of 3-phase PMSM

3. RAS Sensorless Control

In general the RAS algorithm is based on the comparison between the outputs of two estimators. The error between the estimated quantities obtained by the two models is used to drive a suitable adaptation mechanism which generates the estimated rotor speed. The RAS algorithm is well-known in the sensorless control of an induction motor, and has been proved to be effective and physically clear [5-7]. The RAS proposed in this paper is using the state observer model with the current error feedback and the magnet flux model as two models for the back-EMF estimation.

A. State observer configuration

Here, the estimated currents may be replaced by the measured currents, and the order of the observed states may be reduced. This paper uses the reduced order observer.

Fig. 2 shows the block diagram of the reduced order state observer for the back-EMF estimation.

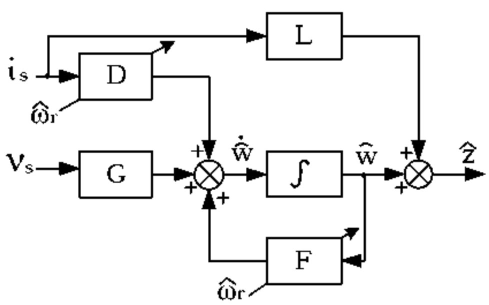


Fig. 2 Block diagram of the reduced order state observer

B. RAS configuration

This paper proposes a sensorless control algorithm based on the RAS for the speed sensorless control of a PMSM. The proposed RAS is using the state observermodel of (17) and (18) and the magnet flux model of (9) and (10) as two models for the back-EMF estimation. The rotor speed is generated from the adaptation mechanism using the error between the estimated quantities obtained by the two models. The proposed RAS algorithm has a robust performance through combining the state observer model and the magnet flux model.

The overall system of the proposed sensorless control algorithm is shown in Fig. 3.

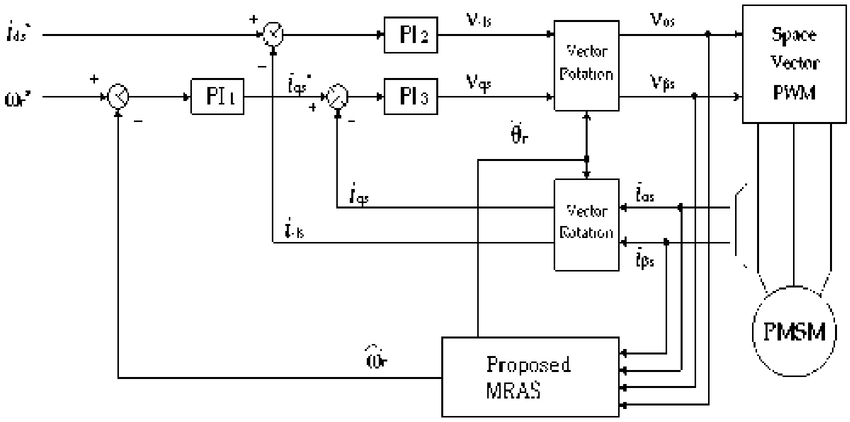


Fig. 3 Configuration of the overall system

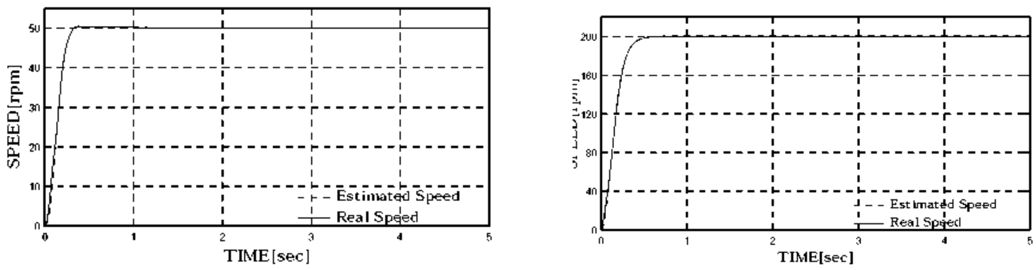
4. Simulation

The simulation has been performed for the verification of the proposed control algorithm. Table 1 shows the specification of the permanent magnet synchronous motor used in the simulation and experiment.

Table 1. Motor specification

Number of poles	8	R_e	1.2 Ω
Nominal current	5.5 A	L_e	4.23 mH
Nominal power	800 W	K_e	0.143 V sec/rad

Fig. 4 (a) and (b) show the speed responses in the speed commands of 50rpm and 200rpm and in the no load. As shown in Fig. 5, the proposed sensorless control algorithm has good speed responses in the low and high speeds.



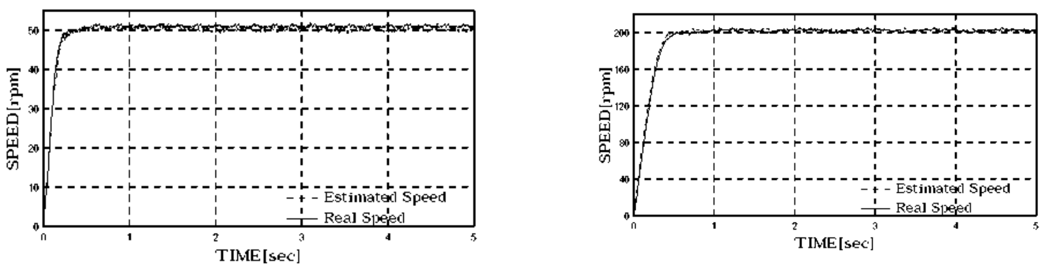
a) b)

Fig. 4 Speed responses in the speed command of (a) 50 rpm and (b) 200 rpm

5. Experiments and Discussions

The experiments have been performed for the verification of the proposed control algorithm. The microprocessor system (80586/150MHz) is used for the digital processing of the proposed algorithm.

Fig. 5 (a) and (b) show the experimental speed responses in the speed commands of 50rpm and 200rpm and in the no load. The proposed sensorless control algorithm has good speed responses in the low and high speeds same as the simulation result.



a) b)

Fig. 5 Experimental speed responses in the speed command of (a) 50 rpm and (b) 200 rpm

6. Conclusions

This paper proposed a novel speed sensorless control algorithm of a permanent magnet synchronous motor. The proposed control algorithm is aided by a Reference Adaptive System using the state observer model and the magnet flux model as two models for the back-EMF estimation. The rotor speed is generated from the adaptation mechanism using the error between the estimated quantities obtained by the two models.

The simulation and experimental results indicate that the proposed algorithm shows good speed responses in the low and high speeds, and shows robust speed responses in the stator resistance and back-EMF variations. Especially, the proposed algorithm shows a better performance in the parameter variation compared to the conventional algorithm.

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ELECTRIC DRIVES WITH FIELD ORIENTATED CONTROL FOR NAVAL MECHANISMS

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Abstract: *In this paper, the authors propose the embracing of a solution regarding the implementation of the direct vectorial control in torque and flux on the induction motor with squirrel cage rotor for the electrical drive systems of the naval mechanisms. Through this is desired the replacement of the actual technical solution, at which for the drive of the loading equipments are used asynchronous motors with squirrel cage rotor with three speed steps obtained with the help of three separate statoric windings in star connection. The main disadvantages of the actual solution are: large size, complicated construction of the driving machine, the modification of the speed can be made only in steps. In order to eliminate these disadvantages a technical solution is proposed which keeps the induction machine as the executive element, but at which the control of the movement, that assumes the control of the speed and/or the control of the position respectively the control of the torque is made through direct vectorial control in torque and flux. The main advantages that are obtained are: - a fast answer in torque and functioning in a wide range of speed, a robust control and relatively simple to implement, does not need current regulators and coordinate transformers, does not need a decoupling circuit of the statoric voltage equation and neither a separate vectorial modulator for the command of the PWM inverter, an efficient rejection of the disturbances is assured and it folds well on the numerical control, a normal asynchronous motor with squirrel cage rotor can be used, with only one winding on the stator, the modification of the speed being made through vectorial control, the command panel is eliminated, on which the direction, acceleration, breaking contacts and time relays can be found.*

Keywords: *Naval mechanisms, electrical drives, induction machine, ABB inverter, acquisition boards, vectorial control, lifting equipment*

1. Presentation of the proposed solution

In this paper a technical solution is proposed which keeps the induction machine as the executive element, but at which the control of the movement, that assumes the control of the speed and/or the control of the position respectively the control of the torque is made through direct vectorial control in torque and flux. The classic direct vectorial control in torque and flux implemented at the study of the naval drive mechanisms assures the direct control of the statoric flux and of the electro-magnetical torque through selecting the optimum way for switching of the PWM inverter with the IGBT transistors. In this way the switching is realized so that the flux and torque error to be enclosed in hysteresis band with the well determined goal to obtain a fast answer in torque and also to reduce the switching frequency of the inverter.

The following advantages are obtained:

- a fast answer in torque and functioning in a wide range of speed is assured;
- the solution is robust, relatively easy to implement and does not need current regulators and coordinate transformers;
- does not need a decoupling circuit of the statoric voltage equation and neither a separate vectorial modulator for the command of the PWM inverter;
- an efficient rejection of the disturbances is assured and it folds well on the numerical control.

The implementation of the direct vectorial control in torque and flux for the electrical drive systems with asynchronous motors for the naval mechanisms was implemented by the authors first on a laboratory model, and then on a naval lifting mechanism on the Albatros ship. The main experimental results are presented in the following section.

2. Data acquisition

The extension of numerical measurements is linked with the increasing of the measurement precision and also with the possibility of numerical processing of the signals and was possible as a result of the progress made in the realization of the integrated circuits, which offers:

- increasing of the complexity and liability of the circuits;
- the realization of components with very similar parameters (for the resistors, smaller differences that 1%, for capacitors smaller differences than 0.1%);
- the measurement of the time (frequency).

In order to highlight the conversion and command processes in the present drive, three signals were acquired: the feeding voltage of the motors (inverter output voltage), feeding current of the motor (inverter output current) and the speed. For this, two acquisition boards were used connected in series, according to the scheme from Figure 1.

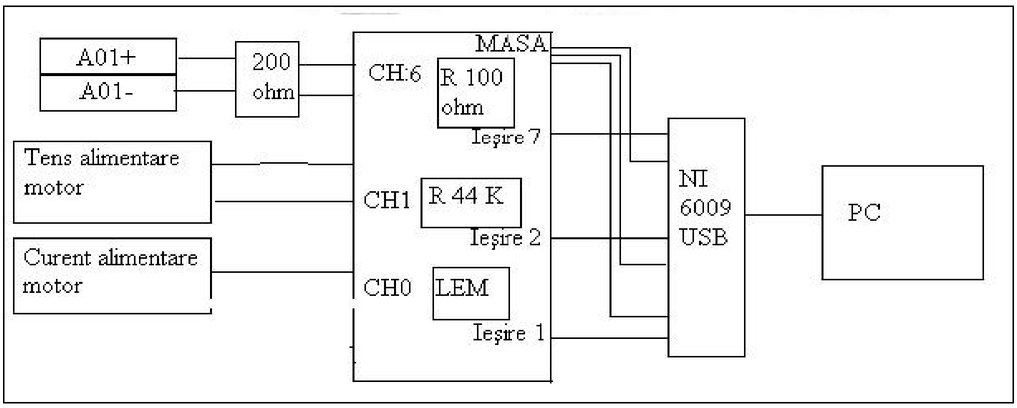

Figure 1 The acquisition scheme

From the output terminal of the inverter a phase is taken and led through the LEM. The role of this is a current sensor, at the output of the acquisition board exists a voltage proportional with the current. On any two phases of the output voltage of the inverter are connected the two wires which collects the voltage signal. This is in parallel with a 44 kΩ resistor, at the output the voltage being proportional with the input voltage but a much lower value that can be supported by the second acquisition board. The converter has two outputs in current proportional with the speed and torque. For both the interval is from 0 to 20 mA, corresponding to a speed of 0 to 100 % from the nominal speed written as parameter in the ABB equipment. Since the acquisition board does not measure small currents, a high precision resistor of 200 Ω is put in parallel with the outputs. Like this, the voltage is applied on the input of the acquisition board on the 1,8 kΩ resistor. At the output of the first acquisition board results a voltage proportional with the speed. From the scheme of the acquisition board presented in Figure 2 and 3, it can be observed which channel corresponds to each acquired parameter. So, the feeding current of the motor enters on channel 0 and the output signal is acquired on the 1-11 pins. The feeding voltage of the motor enters on channel 6 heaving the output signal on the 7-11 pins. The speed enters as a voltage on channel 1, the output from the acquisition board being made on the pins 2-11.

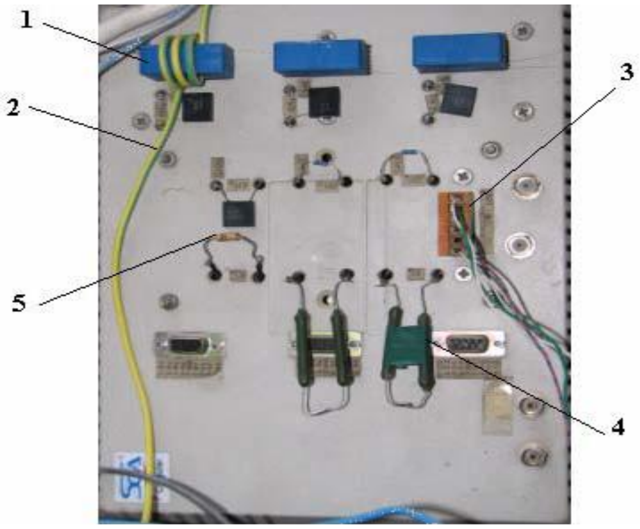

Figure 2 Signal conversion board from the invertor:
1- current LEM; 2- the used phase for the acquisition of the current from the output of the inverter; 3- inputs used for voltage aqusition;4 – 44kΩ resistor corresponding to the feeding voltage of the motor;
5- 1,8kΩ resistance, corresponding to the speed signal.

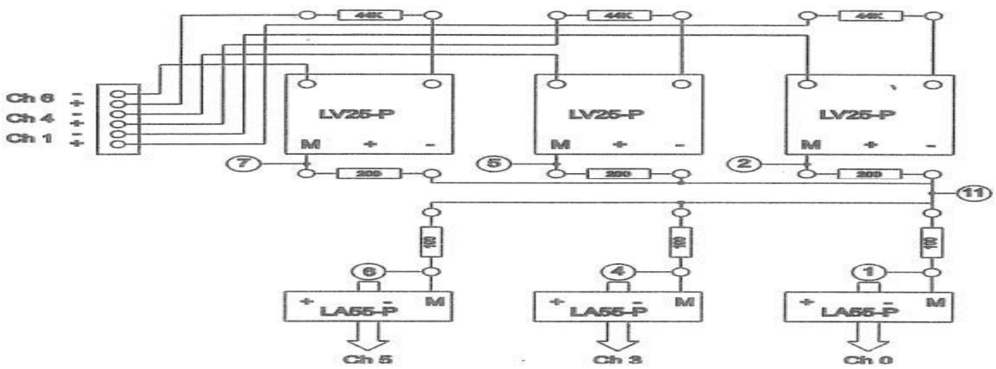

Figure 3 The acquisition board scheme.

The second acquisition board is made by National Instruments (see figure 4). It is a poor precision board, low cost, thought mainly for laboratory applications and teaching purposes. The benefit of this board is that it can be connected through USB port to the computer, has 12 digital inputs. The connectors can be very easily changed. The transfer speed can be up to 48 Kbit/sec. The range of the input voltages is ± 1 to 20 V and of the output voltage between 0 and 5 V.

Figure 4 National Instruments 6009 acquisition boar

To process the data sent by the second board, a Matlab program is used, conceived by the research comity, in order to visualize the signals. The simultaneous and real time display of the three signals is obtained. The maximum time length in which the acquisition can be made is 1 sec. This restriction is given by the construction of the NI 6009 acquisition board. For the speed signal of the speed in normal functioning regime a filtering is made. In dynamic regime this filtering is aborted in order to observe better the increase and decrease of the speed. In the following are presented the obtained waveforms.

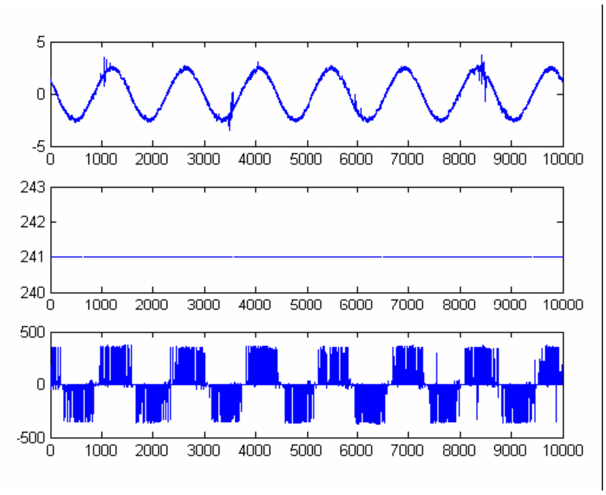

Figure 5 The normal functioning regime at 20 % of the prescribed speed

On the third chart the frequency of the voltage signal of the motor can be observed. Also on this signal can be observed the results after the “cutting” the continuous voltage from the intermediate circuit. The signal corresponding to the speed, the chart in the center, is filtered through the realized Matlab program. An arithmetic average is made of all of the acquired values on the signal of the speed, thus a straight line is obtained that is parallel with the time axis. In the first chart the signal corresponding to the current on a feeding phase of the motor.

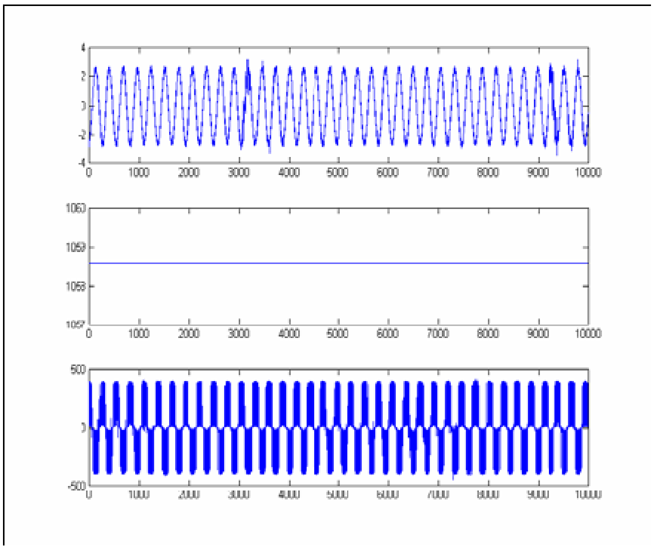

Figure 6 The normal functioning regime at 90 % of the prescribed speed

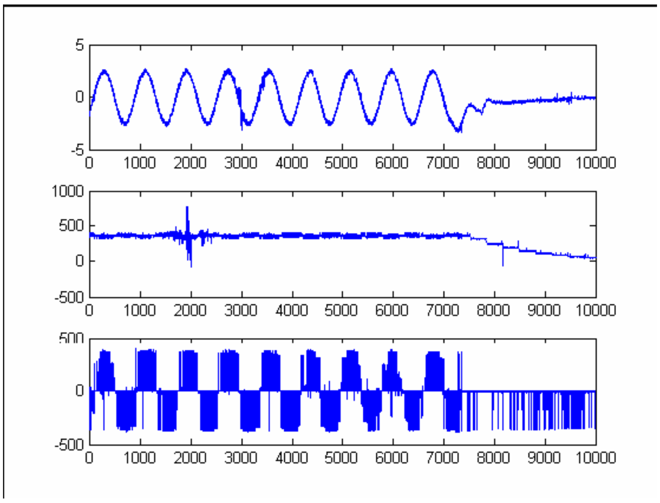

Figure 7 Stopping at 40 % of the prescribed speed

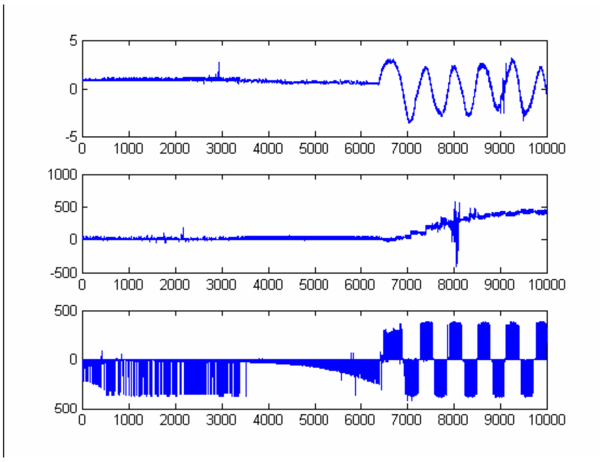

Figure. Starting regime until 40 % of the prescribed speed

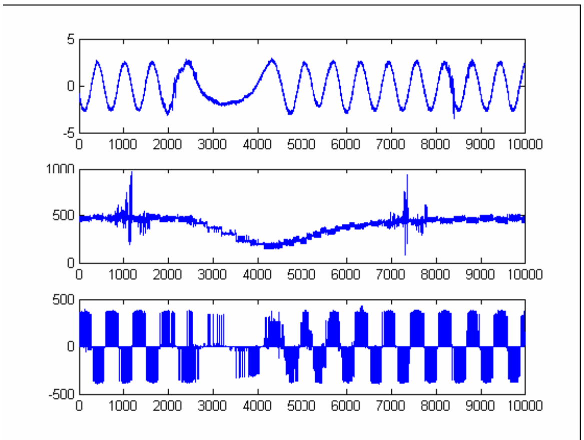

Figure. Direction change

4 Conclusions

The practical implementation of the purposed solution highlights the following advantages:

- there is the possibility of the adjustment of the speed of the motor at predefined values before the starting;
- the steps of the speed are commanded through the closing or opening of some contacts; the potentiometer mounted on the controller is not used;
- the command being external, can be mounted in separate command room or in a place where the operator is protected;
- the usage of the converter commanded in voltage reduces the power consumption. More than this, it presents visible advantages regarding the safety of the operations. Also, the compact form of this devices optimizes the used space in the distribution panel;
- has a high degree of precision and very good dynamic in the conditions of a variable load and speed;
- there exists the possibility of a permanent display of the output parameters from the inverter.

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THE ELECTRIC CHARGE ACCUMULATION TO OIL TANKERS

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Abstract: *The paper makes an analysis from an electrostatic field point of view of installations onboard oil carriers leading to the production and storage of electrical charges. We take a close look at the loading gear of specialized ships carrying oil products starting with steam operated pumps and ending with metallic pipes and filters. When cleansing oil tanks with oil, water, and then steam, washing plants are another source of electrical charges. The inert gas plant may generate an electrical charge of more than 0,6 $\mu C/m^3$ due to the fact that the inert gas has a complex composition and it goes through the exhaustion nozzle in its way to the oil storage compartment. The drain arrangement collecting the water, oil and flammable products residues through piston pumps, pipes and collecting tanks constitute another source of electrical charges. Taking all these installations into account we have attempted to make an evaluation of the electrostatic energy stored within oil products onboard a specialized ship.*

Keywords: *electrostatic field; bilge; heating device; washing installation; steam pumps*

1 Introduction

When an electrostatic discharge occurs onboard a ship specialized in carrying oil products we notice the following sequences: the movement of the fluid triggers the occurrence of electric charges which generate an electric current; the electric charges generated thereby are accumulated in a low conductivity environment; the accumulated charges produce an electric field; due to the electric field there appear electric discharges which may have enough energy in order to light up the fuel.
If one of these sequences is not fulfilled then the ignition doesn't take place and consequently only by knowing the substance can we take safety measures onboard.

The electric charge of oil products may appear by ionization – as a result of rubbing, by polarizing molecules – following the anisotropy of Van der Waals forces (at the free surface or at the contact surface) and due to the washing procedures of cargo compartments.

We can account for the occurrence of electric charges by the process of electron transfer between molecules of the oil product and of the impurities contained by the latter [2]. In the case of cargo compartments the amount of accumulated electric charge raises proportionally to the load volume, the surface contact to the metallic walls and the surface of the liquid mirror. [3]

In the case of transport pipe system, the amount of electric charge has been analyzed depending on the flow, the type of liquid stream (continual, jet, drops); on the period of time during which there is oil fluid in the system, on the electric conductivity of the hydro carbonic fluid, on the existence of additives.

During lab experiments, in the case of negatively charged fuel we have noticed two types of electrostatic discharges. The former consisted of a bright channel which appeared on the discharge trajectory and then spread over the liquid mirror. The latter occurred on the form of a spark. Both discharges have been permanently accompanied by the apparition of Taylor cones –small conical distortions at the surface of the oil product due to the distribution of the electric field. In the case of positively charged fuel we have noticed only electrostatic discharges by Corona effect, of low energy, in restricted areas, accompanied by background noise. [4]

Discharges by spark are not stable; the electric current tends to rise infinitely and the discharge is very swift. The discharge current by Corona effect is auto stable.

2 The procedures of inertia and washing-up of cargo tankers

The evacuation gas from the propulsion engines or the gas generators which contain within themselves carbon dioxide mainly, are cooled off and cleaned from soot and sulfur dioxide with the help of sea water in the depurator and then distributed in the cargo compartments through a system of pipes with the help of centrifugal ventilators. Despite the fact that these gases are often used as inert substances, they may become electro statically dangerous as the gas jet acquires an electric charge of above 0,6 $\mu C/m^3$.

Within a hydrocarbon liquid, the molecules interaction with one another by Wan der Waals and Coulomb forces. Only one component of the force that is perpendicularly directed on the surface of the liquid operates on the molecules to be found on the surface of the liquid. If the oil product comes in touch with the inert gas then this force is directed inside the cargo compartment. If the contact with the oil product is done With the compartment wall or with another liquid, this force may be directed inside or outside the liquid, but it will always preserve its perpendicularity to the separating surface of the phases.

The cleaning operation of cargo tanks triggers a constant accumulation and separation of electric charges. The cleaning is achieved by means of oil or other combustible products being transported, followed by a second washing with water and steam and in the end there is ventilation and gas free. If the cleaning procedure is not performed appropriately there will be stratified residues on the compartment walls due to the high density of the combustible gases. As for a cleaning procedure, electrostatic discharges depend on: the ratio between electric charge generation and the stress relief, the potential gradients which arise during the process, the volatility and inflammability of the oil products or of the fuel-air-water mixture, the minimum ignition energy of the mixture in the cargo compartment.

In order to prevent the occurrence of electrostatic discharges with washing devices onboard ships we can install an electric stress-relief screen around the washing nozzle or a set of conductors placed in a wreath around the same nozzle, or a conductor wire in the center of the cargo compartment between the sky and the bottom.

The existence of the spherical concave screen around the nozzle reduces 6 times the electric field as compared to the case when the screen is missing. When using the stress relief system with conductors, at a density of 0,6 $\mu C/m^3$ in the oil fluid leads to a 4 times reduction in the electric field unlike the case when these points lack. The presence of the conductive wire has reduced 4 times the maximal electric field to the top of the nozzle from 900 KV/m to 200 KV/m. Both for the spherical screen and for the stress-relief points, it is necessary that these be made of materials with a resistivity high enough to limit the energy corresponding to the electric discharges at lower values than the ignition ones.

The existence of metallic wires leads to the apparition of flowing currents to the hull of the ship. These currents lead to a decrease in the load accumulated in the mass of the oil fluid. For a cargo compartment with a height of 15 m for a ship of 65 000 dwt, the flow current is of 2·10–5A, which supposes a relaxation of the electric charge with a time constant equal to 10 seconds that ensures am adequate stress relief. [4]

The presence of water drops and other impurities in suspension within oil products is a source of electric charges. When water washing of the cargo compartment occurs – a polar-heavy material- we have noticed that on the surface of the oil product-water mixture there is a layer of oriented dipoles. These dipoles form a superficial field, which depends on the dielectric constant. The water molecules orient themselves in the electric field and form bridges along which electrostatic discharge is favoured. The parameter, which regulates the production of electrostatic discharge, is the dielectric stiffness of the oil product, fuel vapours, water and inert gas on the free surface of the cargo compartment.

3 Unloading a slop tank and the role of steam coils in the cargo compartment

Residue tanks must be designed so that the positioning of exists, the baffle plates and waste weirs, as well as the way of draining away the collected products should be achieved in such a manner that there be no turbulence or accumulation of electric charges inside the emulsion of water and oil products.

The unloading pipes of the residue tank must be positioned as near to the lower part and the unloading rate must decrease with the pumping –out of the water and the leaving-over of oil residues. This is done with a view to minimizing the accidental mix with the water that is left and to decreasing the spraying and the mist. In the residue tank the surface potential must stay below 35 kV, the volume density of the electric load must be under 0,1 μC/m3 and the intensity of the electric field should be below 28 KV/m [4].

During transportation onboard specialized ships, oil products are kept at a quasi-constant temperature by means of the steam circulating inside the coil which goes through the cargo tanks. Breaking the coils triggers an increase in the quantity of accumulated load in the hydrocarbon fluid.

4 Mathematical relations

The relation between the maximum electric density on the surface of the oil product and the minimum ignition energy is

$$\varsigma_v = 9.10^{-10} \frac{W_{\min}^{\frac{1}{4}} . e^{-\frac{k}{4}}}{S.\Delta\varphi} \tag{1}$$

Where W_{min} - the minimum ignition energy; $W_{min} \geq 10^{-4}$J

$\Delta\varphi$ -the surface potential

S -the surface of the oil product mirror

$$\Delta\varphi = -\frac{(1-\beta)}{\varepsilon}(\varsigma_V + \beta.\tau.\frac{kT}{q.z.\Delta m}) \qquad (2)$$

$\beta = \frac{Dc-Da}{Dc+Da}$, Ds Da - diffusion coefficient of the cat ions, of the anions respectively

K- Boltzman's constant

q - Electric load

Δm - Diffusion coefficient

τ - Conductivity of the oil product

ε - - Electric permittivity of the oil product

T - - Temperature

z - -Valence of the substance

$$\varsigma_V = \frac{\chi.v.\Delta_0.L}{\frac{\Pi.d^2c}{4}}$$

$$\chi = f.c.t$$

χ_e -proportionality factor which depends on the molecule concentration (c) Faraday's constant (F) and time (t)

Δ_0 -relative rugousity of the walls

L-the length of the pipe

$A_c = \frac{\Pi dc^2}{4}$ -the area of the transversal section on the cargo pipe

v – the flowing speed of the fuel fluid through the pipes

5 Conclusions

The concentration of the electric charges to the metallic walls of the cargo compartment is on one hand due to the diffusion of the ions in the electrostatic field made up of existing charged particles and, on the other hand, to Van der Waals forces. Van der Waals forces that operate among non-polar molecules from within the liquid appear due to the following physical phenomenon; because of the vacillating movement of the atoms, they display at any time non-void instantaneous electric momentum vectors, which are variable in time but whose temporal average is null.

With the help of the inert gas installation we create a buffer between the oil product and the air. At the surface of the oil product from the cargo compartment there is a oriented dipole layer; consequently there appears a surface potential, of separation between the liquid-vapor and the inert gas. The oriented dipoles make up a superficial field, which depends on the dielectric constant. Under the influence of the oriented dipoles field, the anions that are to be found inside the liquid are attracted to the positive poles of the dipoles making up at the separating surface a negative cloud of the absorption layer. A second cloud of the double layer is formed by cations, which screen out the charge of the anions. The bound of the anions with the dipoles is strong and the cloud remains unmarked. The bound of the cations is much weaker and they can move about freely in the oil product making up a mobile layer of the double layer. [5]

The electric charge which appears and accumulates between the surface of the oil product in the cargo compartment and its sky (ceiling) in the space of inert gas, create an electric potential which may lead to electrostatic discharge. These discharges may occur between the gear used to bled combustible gas in the cargo tank and the grounding system of the ship –sparks- or between the gas and the grounding perturbations (when cleaning) - Corona.

The need to have a thorough knowledge of the electrostatic regime in the case of oil tankers is made evident by the fact that it helps us acquire a quantitative measure of the electrostatic conditions which arise and be able to identify the types of electrostatic discharge. In the process of exploiting an oil tanker we must establish the distribution of potential and electric charge by means of calculations and experimental models.

At the interface between a solid surface and a liquid or between a liquid and a gas there appears an electric double layer made of the compact layer and the diffuse layer operated by diffusion, migration, and convection phenomena. [6]

The paper has analyzed the influence of some installations in the accumulation of electric charges with an oil tanker.

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SPECIFIC NAVAL EQUIPMENT

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Abstract: *Control engineering embraces instrumentation, alarm systems, control of machinery and plant previously known under the misnomer of automation.*
Control engineering can be applied not only to propelling and auxiliary machinery but also to electrical installations, refrigeration, cargo handling (especially in tankers) and deck machinery, e.g. Windlass control. Opinion still vary on such matters as the relative merits of pneumatic versus electronic system and whether the control center should be in the engine room or adjacent to the navigating bridge. Arguments against the exclusion of the engineer officer from close contact with the machinery are countered by the fact that electronic systems are based on changes other than those of human response. Automated ships (UMS) operate closer to prescribed standards and therefore operate with greater efficiency. The balance between the possible and the necessary would be achieved in this case by combining automatic monitoring of all the likely fault conditions, with routine machinery space inspection say twice a day [1…9].

Keywords: *naval equipment, automation, machinery, system*

Planning the system

Planning of the automation system, by which is meant the total complex of remote and automatic controls and plant instrumentation must take account of several basic parameters:

1. The intended service of the ship.
2. The intended manning arrangements.
3. The type of propelling machinery.
4. Ship maintenance policy.
5. Classification society and notation required.
6. Ship resale value.

The above list of "design inputs" is by no means complete, but represents the major factors, which should influence the design of the automation system.

Experience has shown that where there has been some failure to achieve all that was expected it is largely due to lack of planning. Successful planning involves integrating and coordinating the system, as a whole and this cannot be achieved if sections are in different hands. Haphazard methods by independent concerns have resulted in conflicting and unworkable systems. For example, sensors have been used at the instigation of one interested party and without consultation with, for example, the supplier of the computer or the data-logger only to find later that the output is incompatible.

It is also essential the control engineer should have practical knowledge and experience of the plant to be controlled and that the plant supplier should concur regarding facilities for accommodating and positioning the sensors.

A procedure, which has been advocated for ensuring success, is that the ship-owner should, at the outset, state in broad terms what he requires. The shipbuilder should then prepare an outline specification to meet the owner's requirements and from this the control engineer can prepare a detailed specification. All three parties should then get together and agree the control specification. Hitherto there has been too little feedback information and experience from the ship but control engineers and ship-owners are now appreciating that this is important. If owners or builders have preferences for any particular make of component for any particular make of component; it is at the planning stage that agreement should be reached.

The owner will need to consider operational and economic issues to decide how far to go and what financial benefits he can expect from each section. For example, in a refrigerating plant, push-button starting from the control console may not be justified, as it is an infrequent operation, which can be performed manually, and so centralization can be confined to instrumentation and alarms. The essential factors for successful systems are:

1. Reliability.
2. Simplicity.
3. Ease of operation and maintenance.
4. Suitability for marine conditions.
5. Facilities for servicing (especially in foreign ports).

Marine conditions involve ambient temperatures, humidity, vibration and saline atmospheres but also the physical conditions inevitable during construction, installation and trials. These apply to all parts of the system-sensors, instruments, consoles, computers, etc. Paint spraying, asbestos lagging, welding, staging and dirty surroundings can play havoc. Fitters and erectors have no respect for such equipment and many sensors have served as a footstep.

Systems must embody "fail safe" features and this aspect must be studied analytically in the planning stage. All possible sources of failure and their consequences must be covered. For example, if a fuel injection system is such that a spring is balanced by fluid pressure acting on a piston then loss of fluid may result in full fuel admission to the engine and a dangerous condition exists. The arrangements must ensure that failure of the controlling medium will result in either the speed remaining constant or that is reduced.

Fail-safe principles can be interpreted in different ways, such as complete stoppage of an operation or reverting to some other (safe) state. In suitable cases it can mean "fail-as-set", i.e. continue as at the time of failure, sometimes referred to as "failed-as-is".

A vital part of planning procedure is planning the pre-commissioning trials and calibration. This must be considered and agreed by the builder at an early stage so that he can include it in his overall program and delivery date and, when the time comes, provide the essential facilities.

It is not unusual for a comprehensive system to include 300-400 control points widely distributed and each requiring individual checking for operation and possibly calibration. This is time consuming and can only be done when installation is complete and ship's services are available. It cannot be postponed until after the sea trails. A detailed test program and timetable, agreed by the shipbuilder is therefore essential. Whit all systems there is an initial period of teething troubles and these must be tracked down as far as possible before the sea trials. This applies particularly to closed-loop systems.

Simulators can be provided in some cases, which make possible to test the entire electronic equipment by providing similar responses to those anticipated under service conditions. They can form part of the permanent installation so that, for example, prior to arrival in port, the navigating officer can himself simulate operation of the engine telegraph.

Sensors

Sensors play an essential role in all systems for transmitting information to control and other remote positions. The quantities necessary to sense include counting, fluid flow, humidity, liquid levels, noise, position, pressure, salinity, smoke density, speed, strain, temperature, viscosity, torque, power, etc.

The type of sensor must take into account the relative importance of the effect of its presence on the quantity to be measured, together with the extraneous effects by or on the sensor. For example:

1. It should not effect the quantity to be measured, e.g. flow metering.
2. The effect of ambient and adjacent temperatures should be either known or be capable of elimination.
3. Speed of response in respect to rapid changes.
4. Independence from magnetic fields, humidity, barometric pressure, local heat.
5. Independence from variations of electrical supplies (e.g. frequency and voltage) or be provided with means for compensating for variations.
6. Linearity, hysteresis, repeatability and zero-point drift are also important.

Sensors may be required to initiate mechanical operation, for example, such as the high forces required to operate cargo valves in tankers and for hatch closing and opening and as most sensors cannot provide the mechanical effort required this can be provided via transducers. The electrical or pneumatic signals obtained from them can in turn operate alarms, relays or instruments. Bourdon tubes, diaphragms and floats can provide sufficient power to operate instruments directly or can act as transducers.

Measurement of process conditions

The range of parameters to be measured in merchant ships includes temperatures, pressures, level, and speed of rotation, flow, electrical quantities and chemical qualities. Instrumentation used for remote information gathering purposes invariably converts the measured parameter to an electrical signal which may be used to indicate the measured value on a suitably calibrated scale, provide input information to a data logger or computer, initiate an alarm or provide a signal for process controller – fig. 1.

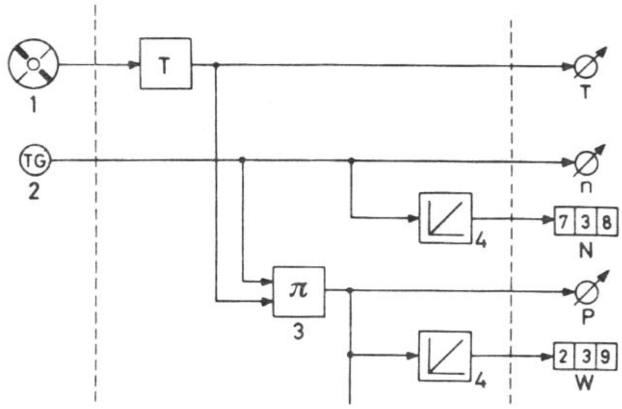

Figure 1 Torductor power meter

1. Torque transducer	*4. Integrator*	*N. Total number of revolutions*
2. Tacho-generator	*T. Torque*	*P. Shaft horsepower*
3. Multiplier	*n. Rate of rotation*	*W. Total power output*

As stated earlier however the more favored means of providing process control information (as opposed to information display only) is to use a pneumatic system.

Controls for generators

In unattended machinery installations it is necessary to provide certain control facilities for the electrical generating plant. These may vary from simple load sharing and automating starting of the emergency generator, to a fully comprehensive system in which generators are started and stopped in accordance with variations in load demand.

Medium speed propulsion plants normally use all diesels generating plant. Turbine ships obviously use some of the high quality steam generated in the main boilers in condensing or backpressure turbo generators, with a diesel generator for harbor use. The usual arrangement on large-bore diesel propulsion systems is a turbo generator employing steam generated in a waste-heat boiler, plus diesel generator for maneuvering, port duty, and periods of high electrical demand.

Diesel generators the extent of automation can range from simple fault protection with automatic shutdown for lubricating oil failure, to fully automatic operation. For the latter case the functions to be carried out are:

- Preparation for engine starting.
- Starting and stopping engines according to load demand.
- Synchronization of incoming sets with supply.
- Circuit breaker closure.
- Load sharing between alternators.
- Maintenance of supply frequency and voltage.
- Engine/alternator fault protection.

Preferential tripping of non-essential loads and restoration when sufficient power becomes available.

It is necessary to provide fault protection for lubricating-oil and cooling services, and in a fully automatic system these fault signals can be employed to start a stand-by machine, place it on line, and stop the defective set.

Turbo-generators The starting and shutdown sequences for turbo-generator are more complex than those needed for a diesel-driven set, and fully automatic control is therefore less frequently encountered. However, the control facilities are often centralized in the control room, together with sequence indicator lights to enable the operator to verify each step before proceeding to the next. Interlocks may also be employed to guard against error.

The start up sequence given below is necessarily general, but it illustrates the principal and may be applied to remote manual or automatic control:

- Reset governor trip lever.
- Reset em'cy stop valve.
- Start auxiliary L.O. pump.
- Start circulating pump.
- Apply gland steam.
- Start extraction pump.
- Start air ejectors.
- Open steam valve to run-up turbine.

Where a waste-heat boiler is used to supply steam to a turbo-alternator, control of steam output is normally controlled by a three-way valve in the exhaust uptake, the position of which is regulated in accordance with steam demand. Surplus waste-heat is then diverted to a silencer.

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THE SIMULATION OF THE EFFECTS OF THE RADIANT ELECTROMAGNETIC PERTURBATIONS UPON A DIGITAL ELECTRONIC CIRCUIT REALISED WITH A TYPE RISK MICROCONTROLLER

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Abstract: *The study analyses the effects of some external radiating electromagnetic perturbations upon an electronic device with type RISK microcontroller, based on a simulation technique of induced perturbations by using some fictive perturbation sources on the mass circuits. For this reason there are presented comparative graphics of unperturbed useful signal and perturbed one, using sinusoidal signals, the latest having comparative amplitude with the useful signal and different frequencies, going through 3MHz (radar frequency zone). The designing of the electronic device with microcontroller, as well the performance of the simulation technique, has been made with TINA standard program.*

Keywords: *electromagnetic perturbations, digital electronic circuit, rish microcontroller*

1. Introduction

In condition of increasing the number of complexity, diversity and power of the electric and electronic equipment, especially those of electronic microcomputers and microcontrollers, in the residential and activity areas, civilian and military, the risk of electromagnetic pollution trough an electromagnetic interference mechanism is increasing with all negative effects which appear from here, such as the appearance of errors or even major disturbances or damages when working [1], [2].

It's known the fact that the penetration of electronic equipments by the majority perturbation signals is favorite by the existence of the mass circuits and also of the loop circuit. The effects of this perturbations consists in the modification of useful signals (in form, amplitude, frequency, phase), as well as of the reference potential [3], [4].

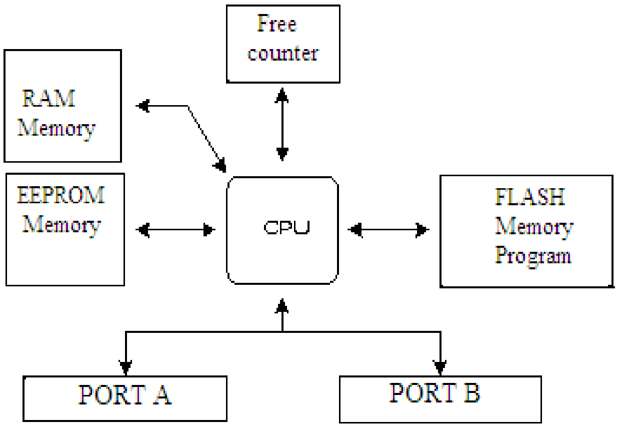
To make evident this phenomenon and its negative effects upon the electronic circuits having the purpose of identification the proper protections measures, in this study are simulated by software perturbation couplings, using a multifunctional electronic device based on PIC16F84 microcontroller. This one belongs to a microcontroller of 8 bits with RISK architecture class and can be used in variety applications, for example in Bank institutions, to verify the smart cards.

2. Short description of PIC16F84 microcontroller

PIC16F84 belongs to a microcontroller of 8 bits class with RISK architecture [8]. Its general structure contains base blocks shown in Figure 1. These one are:

Program memory (FLASH) - for memorize of one written program – can be programmed and erased more than once; this makes the microcontroller adequate for develop of various applications (in our case – for more sort of cards);

EEPROM – memory for dates that need to be saved in case the circuit falls down – it is usually used for important dates, which should not be lost for any occasional damage;



PIC16F84 Microcontroler profile

Fig.1 Structure of the microcontroller

RAM – date memory – it is used by a program and during the execution it contains all intermediary results or temporary data that are not critical for any occasional damage of the supply source;

PORT A and PORT B are physical connections between microcontroller and the outside (connections for user applications). Port A has 5 pines and port B has 8 pines;

FREE RUN is a register of 8 bytes belonging to a microcontroller structure that works without using a program. Every each 4^{th} pulse of the oscillator's clock he is raising the own value until it reaches the maximum of 225; then it starts counting again from zero. It is known that the period of time between two increments of the timer can be used for the measurement of time, being very useful for different applications, including the one already mentioned.

CENTRAL PROCESSING UNIT – realizes the operations given by the program's instructions and it also ensure connectivity between the two blocks of microcontroller.

3. The application's design. The software support

Using TINA program (that fallows the industrial standard SPICE [7]) it was possible the designing of the 3-D image electronic circuits board, containing all the components, and the simulation of the prototype. The results can be seen with the virtual instruments (oscilloscope, selective voltmeter etc.) or diagrams, as the presented application.

Using the integrated module for designing of the circuits, it was made, in case of the electronic circuit, the design of the registered multi-stratification board. This was possible because all used components are stored in data-base software of TINA (respectively – SPICE).

One of the main benefits of the program is the 3-D image of the components and the registered circuit. The figures 2, 3, 4 shows 3-D images of the back-side, of the registered circuit board and of the front-side of the circuit and the figure 5 shows the functional blocks of the device, revealing the inside connections and the outside also.

The most important part of the program is the analysis subprogram. It must be shown the fact that TINA simulator has multiple possibilities of analysis, such as: Fourier analysis, transfer function analysis, analysis of the intern noise of the components, or distortion analysis.

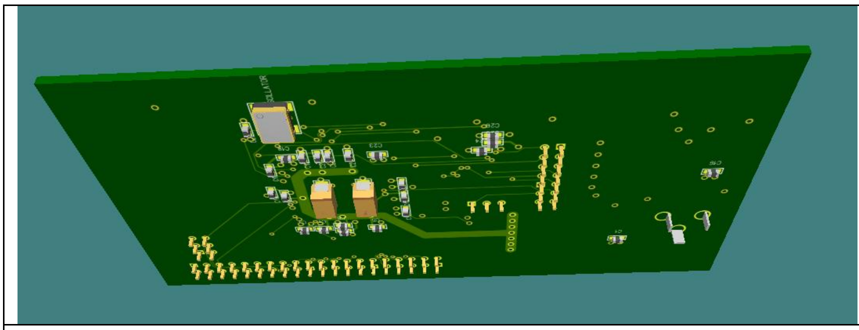


Fig.2. 3-D image of the back-side of the data-base

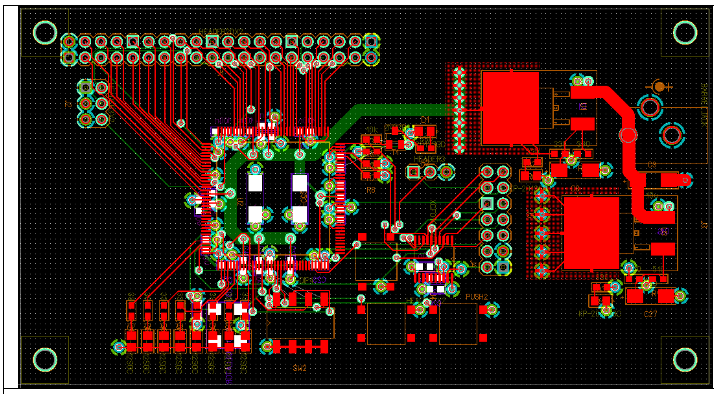


Fig.3. 3-D image of the designed circuit

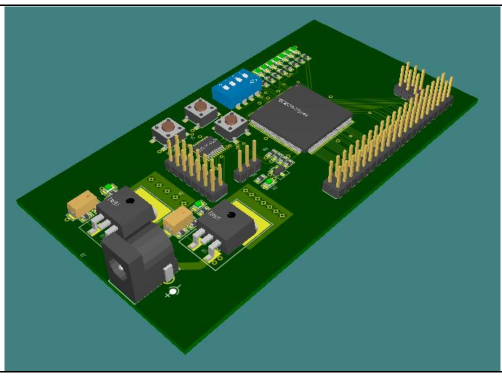


Fig.4. 3-D image of the front-side of the registered circuit

The outputs of this program offer data that can be used later by the subprogram of results presentation; that show this dates in graphics and texts (fig.6, 7, 8, 9, 10 and 11).

4. Simulation possibilities. The simulation of radiant electromagnetic perturbations

The simulation of the device function in different situations offers the possibility of taking into consideration certain terms that refer to:

- identification of effects of parameter's variation of some components;
- study on damages of components from data-base;
- identification of negative effects of electromagnetic perturbations from the environment.

This way it is possible to make determinations on the functionality in static point, on the answer of the circuit at a small signal, on the sensitivity at parameters' variation, on variation of the electronic components` parameters and on their effects consisting in the answer of the simulation of the transitory regime conditions; also on components and electronic circuits function in disadvantageous situations, for example under the effect of the temperature, noise, electromagnetic perturbations or on behavior in frequency (Fourier analysis).

Beside these possibilities the present research adds another one, with a major importance on the security and stability when functioning the microcomputers, microprocessors and microcontrollers, especially on those used in risky process or in institutions as Banks: the simulation of the behavior of this equipments against the radiant electromagnetic perturbations from the environment.

The simulation technique of induced perturbations applied to the analyzed device consists in the simulation of some electromagnetic perturbations propagated by conduction on the mass circuits of the microcontroller, involving the introduction of some fake sources of sinusoidal signal, with amplitude values between 1-4 V, comparable with useful signal, and frequencies of 10 MHz, 100 MHz, 900MHz, 1.8GHz, 3GHz, in radiofrequency and radar scales, (fig. 5 a).

The useful digital signal (output signals, clock signals) having the amplitude of 4 V and the rhythm frequency of 300 kHz, in normal unperturbed condition of the device's functioning, is shown in figure 6. This signal is obtained at P80, P81, P82, P83, P84, P85, P86, P87 pines of the microcontroller.

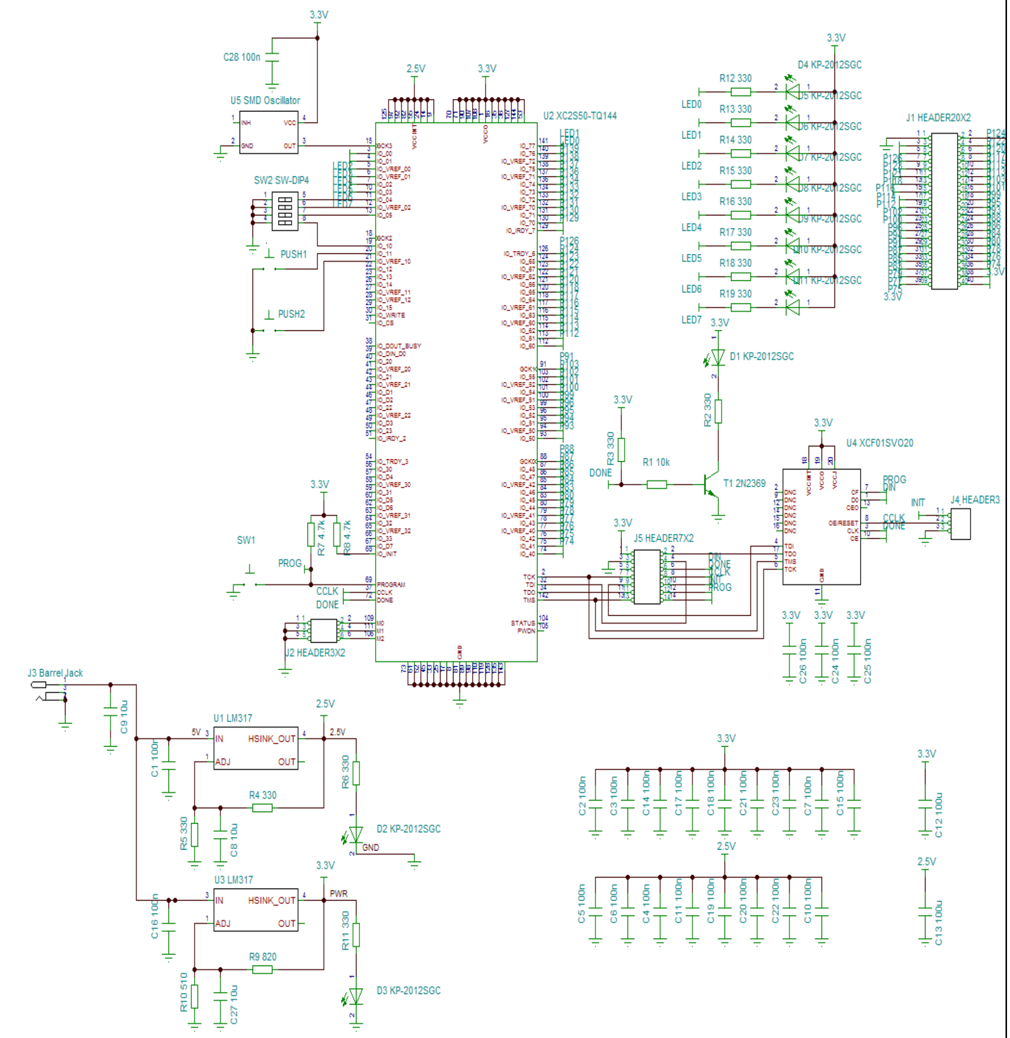


Fig.5 The connection schema of the circuit

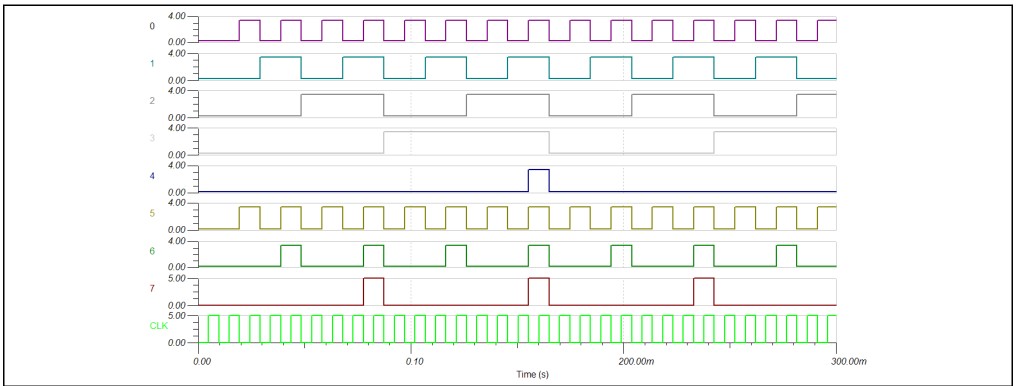


Fig.6 The analyzed signals set in the normal functioning regime (unperturbed)

In fig. 7, 8, 9, 10 and 11 is presented, in opposition, the set of unperturbed and perturbed signals, found at the same pines, after the simulations of some perturbations propagated by conduction, having sinusoidal form, 1-4 V amplitude and 1MHz, 100MHz, 900Mhz, 1.8MHz, 3MHz frequencies on the mass circuit of the microcontroller.

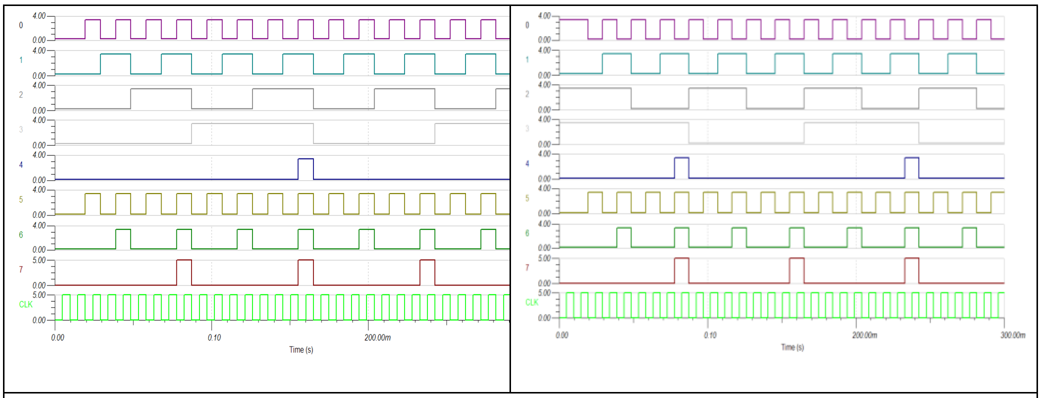


Fig.7 The unperturbed signal (a) and the perturbed one (b) of 1 V, 10 MHz

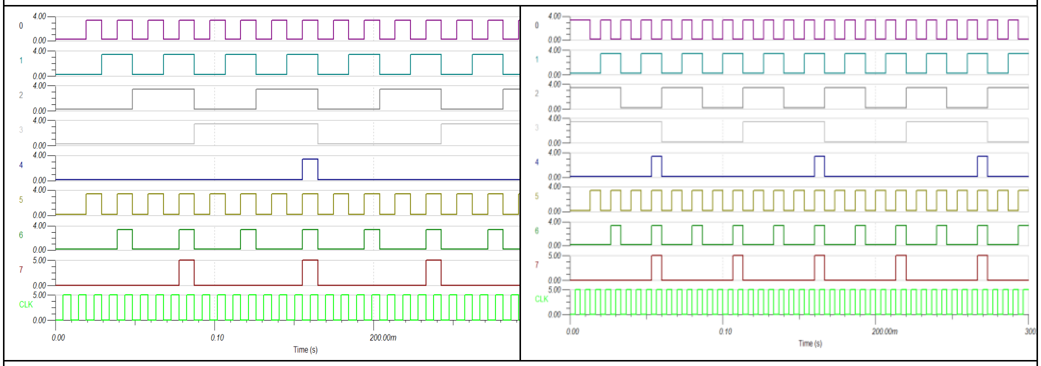


Fig.8 The unperturbed signal (a) and the perturbed one (b) of 1 V, 100 MHz

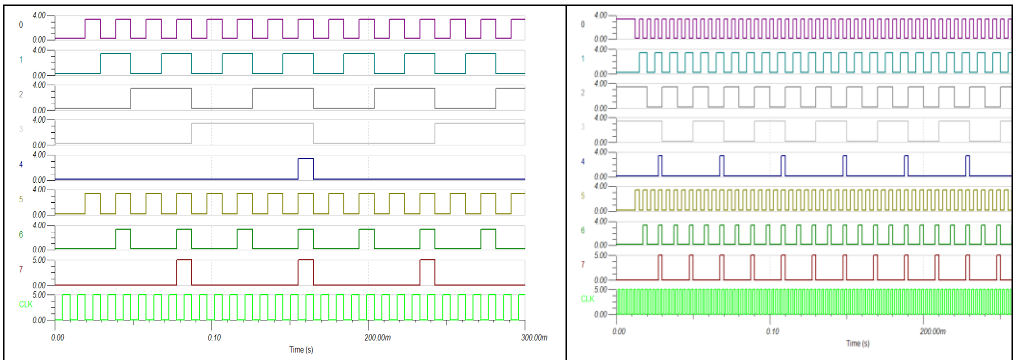


Fig.9 The unperturbed signal (a) and the perturbed one(b) of 1 V, 900 MHz

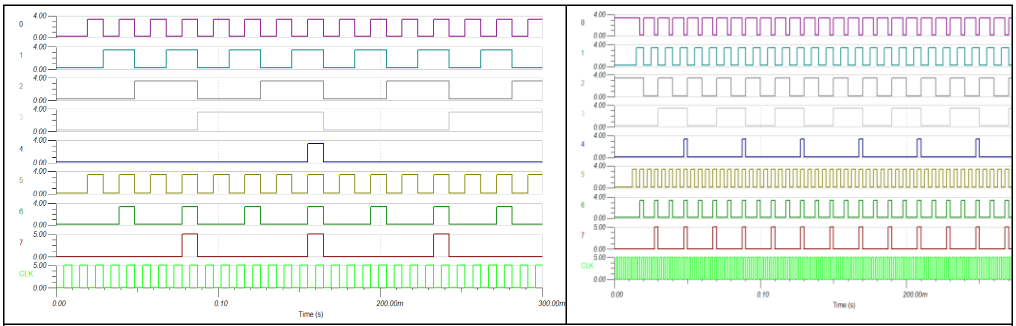


Fig.10 The unperturbed signal (a) and the perturbed one(b) of 1 V, 1.8 GHz

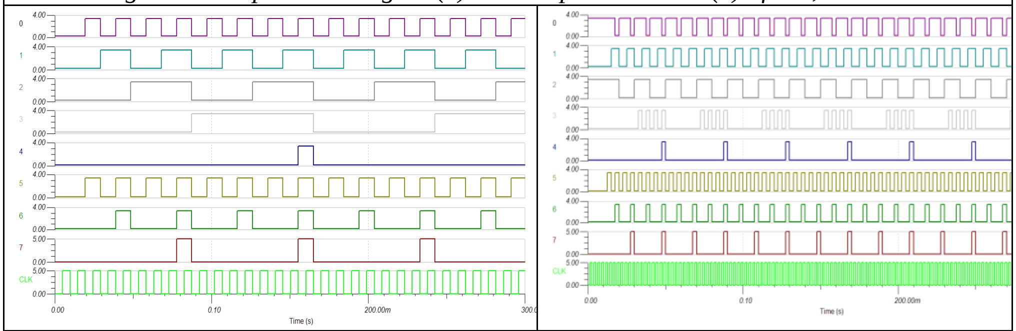


Fig.11 Unperturbed signal (a) and perturbed (b) of 1 V, 3 GHz

Analyzing in comparison fig. 7, 8, 9, 10 and 11 it is notable that as the frequency of the perturbation signal rises and the difference between the frequencies of unperturbed signal and the perturbed one is bigger, the useful signal suffers changes, leading to an inadequate functionality for the device. The same fact can be note concerning the microcontroller clock's signal, (300 kHz in a normally condition).

Conclusions

The simulation shows that the useful signal can be modified by electromagnetic perturbations introduced into the circuit from the environment, resulting negative effects including damages upon the device. In order to avoid these ones, there are necessary anti-perturbing protection specific measures (for example shielding, filtering i.e.) [5], [6].

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EXPERIMENTAL RESEARCHES OVER THE POSSIBILITY OF SEA WATER PROPULSION BY INDUCED CURRENTS

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Abstract: *The induction naval MHD thruster is in fact a linear electric engine in which the metallic moving drive has been replaced with a pipe filled with sea water. The current induced in water interacts with the progressive magnetic field of the inductor. So that is generated forces that make the water to move through a pipe. The sea water has conductivity 10^7 times smaller than copper and because of this reason the quality factor of the sea water engine is very small. The experiments made with classical linear engines having as a moving drive the sea water and are supplied with a three-phased current with industrial frequency (50Hz) do not succeed to make water to move. The theoretical studies showed that increasing the polar step and the frequency may increase the value of the quality factor until values that make the sea water to move. In his paper, the author shows an experiment that demonstrates the possibility of moving sea water using an induction spinning micro engine supplied with two-phased current of 15 KHz. It is described the construction of the micro - engine and that of the supplying system and the experimental results are interpreted.*

Keywords: *sea water propulsion; induced currents*

1 Introduction

Propulsion by induction has the advantage of dispensing electrodes, so that the electrochemical processes is missing, that the gas bubbles will not exist and so will the corrosion. Another advantage will be that of simplifying the commanding processes. It is enough to adjust one single current, the one that goes through the inductor's coils. On the other way every time the ship's speed changes, the magnetic wave speed must be changed too, modifying its frequency. Otherwise we will have a fall of propulsion efficiency. There are two ways to make induction thrusters: with external flow and with internal flow. In the first case, the inductor's coils are disposed on the ship's hull and acts directly on the sea water. In the second case the whole thruster is introduced in a pipe that is inside the ship. If we compare the internal flow with the external one we will see that the first one can reach electrical efficiency situated between 65-85 %, because the field remains inside the pipe situated in the hull. In the external flow case, the way of construction is simpler but the efficiency is reduced. Even though, the displace of inductors is made on the hull so that the fields are perpendicular, at a given distance from the ship, this will loose its perpendicularity. It will appear electromagnetical field's components that don't contribute to the thrust. The fields that are extending to infinite are easy detectible.

In the past, some experiments were made with inductors of a linear electrical engine, on which the moving drive was a sea water pipe. The results were not concludent because of the small quality factor. [1;6] Because of that, the author made a series of qualitative experiments to demonstrate the possibility of seawater drive with an induction microengine powered with a current of high frequency (15 kHz). At this frequency, according to the theoretical study, engine's quality factor should be big enough so the water drive becomes visible. [1:9]

2 Experimental device

To study seawater drive through induction, the author builds a two-phased induction rotary engine, with a fixed cylindrical tank filled with seawater. The microengine was made from the ferrite recovered from linear TV transformers using a diamond cutter to carve and adhesives to paste. In figure 1 is represented the system of yokes and columns of the device.

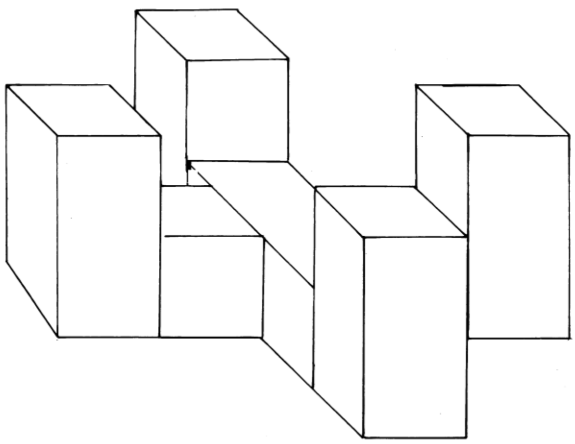

Fig. 1

On the four columns were mounted four coils wrapped in cardboard carcasses. Each coil has 120 spires from a Cu+Em conductor Φ 0,8 mm. On the top of the columns were glued 4 polar parts, and in between was introduced a PVC tank filled with seawater. In the tank water was suspended a teflon piece fitted with 4 blades, with the help of a copper wire Φ 0,1 mm. On top of teflon piece was placed a small mirror. A fine beam of light coming from a projector drops on the mirror and reflects on a measuring rule. In figure 9.2a it is represented a side view, and in 2b a view from the top. [3]

Two opposite coils were serial connected, thus forming two pairs of poles disposed in complanation. The two coils were powered with two approximately sinusoidal currents, out of phase with $\pi/2$. The two currents came from two AF amplifiers of 100 W each. The amplifiers received through a differential transformer two signals coming from a signal AF generator of 15 kHz. One input had a capacitor with a value set so the currents powering the coils to be out of phase with $\pi/2$. Signals level and alteration of phase were tuned, adjusting capacitor the C value and the amplification. The intensities and their alteration of phase were displayed on an oscilloscope with two channels. [8] The block diagram of the device is represented in figure 3.

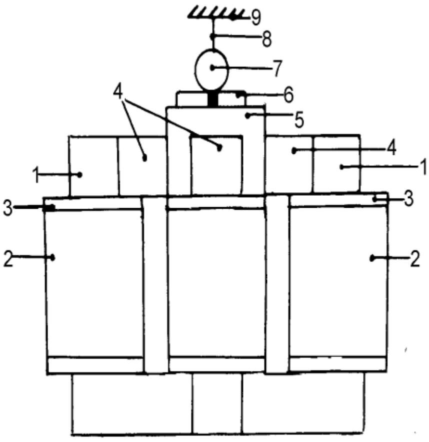

Fig. 2 a

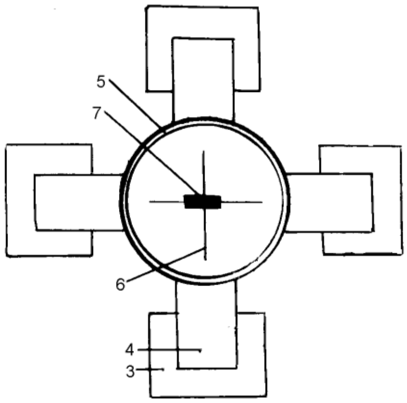

Fig. 2 b

1 - ferrite core 2 - coils 3 - case 4 - polar piece 5 - seawater tank 6 - blade piece 7 - mirror 8 - wire 9 - wire support

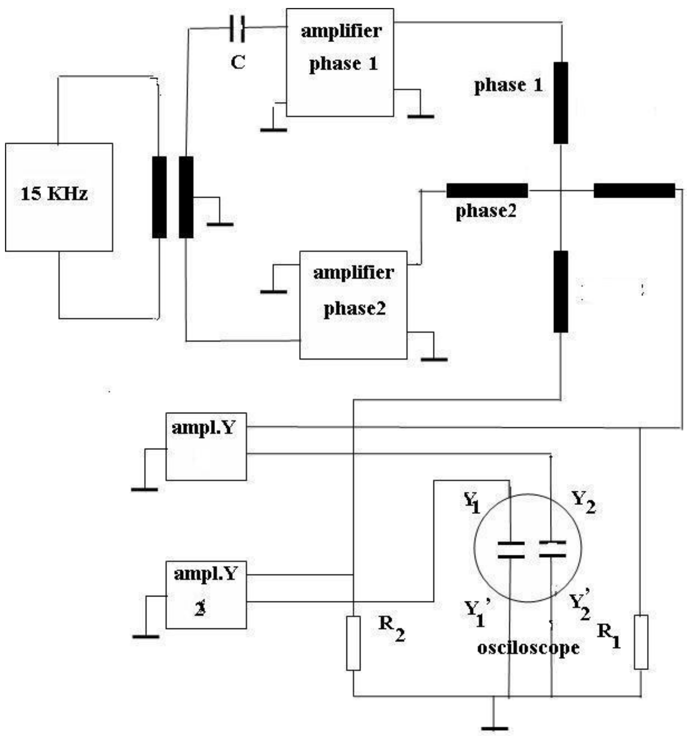

Fig. 3

The voltages from the resistors R1 and R2, each of 0,1 Ω, are in phase with current intensities from the two phases of microengine. The device was powered from a differential generator supplying 2x36 V.

3 Experimental result

In the case when the amplifiers weren't powered, the bright spot on measuring rule oscillate irregularly around neutral position. The amplitude of these oscillations didn't exceed 5 degrees and in author's opinion they were generated by the air currents in the room.

After powering the device, the oscillations kept the same amplitude, but the spot around the oscillations were revolving was drifting away from the initial point. This drift increased in time with the direction of swirling field. This direction was determined by observing the direction of the rotation of a thin aluminum plate cylinder, suspended with a wire between the polar pieces. After turning off the power, the center of oscillation return to its original position in approximately 17 seconds as it is shown in figure 4.

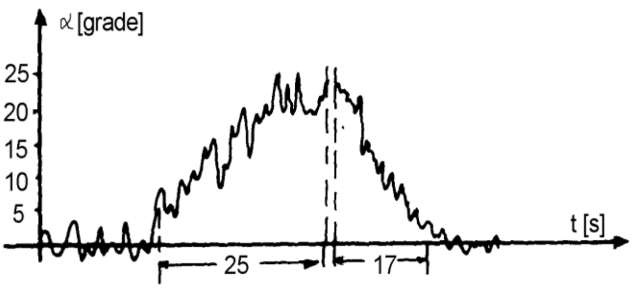

Fig. 4

The author didn't succeed to maintain system power more than 45 seconds due to excessive heating of amplifiers and wiring, and he didn't succeed to observe visually a whirl of water from the tank. Still, the equilibrium position error of the mobile part proves that in the water movements take place which cause this error. The fact that visible macroscopic movements don't appear can be explained with the small value of the forces and with the short period of operation. If the period of operation would be greater, author is convinced that the water driven by the whirling field would be visible, unlike the experiments conducted at 50 Hz. In the future, the author will continue the experiment, building an experimental microengine as a linear engine with a bigger polar step.

4 Conclusions

Building a naval thruster in which the seawater is driven by induced currents, so building a linear engine with seawater as moving drive, has advantages because the system no longer needs the electrodes. This fact represents a major advantage because it was proved that the process takes place at interface between electrode-seawater lead to an efficiency decrease.

The major drawback which stopped until now the making of such a thruster is the fact that the seawater has a very small conductivity in relation with metals and therefore the intensity of induced currents is remarkably small at frequencies of 50 Hz.

A naval induction thruster is in fact an electromagnetic pump, more precisely a linear induction machine with seawater as moving drive.

To have similar performances with a linear machine with solid moving drive, this machine with seawater as moving drive should have a quality factor close as value to the prior. The equality of the two factors presume compensation by some means in the case of machine with liquid moving drive of the handicap resulted from the fact that seawater conductivity is 10^7 smaller than copper.

To compare the two quality factors, the author conducted a study previously upon this quality factor which outlines that this quality factor can be defined as the ratio between active energy dissipated in moving drive and magnetic field energy in air-gap. The value of this ratio differs from the quality factor defined by Laithwaite only with a non-dimensional coefficient, so, according to the author, it can used to characterize the efficiency of a linear electric machine.[1;2;4] Such way to characterize machine efficiency is useful in the case of the machine with liquid moving drive than the classic way in which is defined this quantity. The author calculated the value of the ratio for an induction machine with liquid moving drive using simplifying hypothesis, which, on his opinion, do not influence decisive the results.

From previous equations, it results that if some one wants similar performances for an induction machine with seawater as moving drive like the performances of a machine with solid moving drive, it is necessary that the product between the square of polar step and the frequency for the machine with liquid moving drive to be bigger than the one for the machine with metallic moving drive as many times as the conductivity of the liquid is smaller than the metal. Therefore, a machine with seawater as moving drive having a polar step with a magnitude of meters and working with a current with a frequency with a magnitude of tens of kHz, could achieve similar performances with a classic machine with metal moving drive.

Making polar steps that big and operating at such high frequency create some technological problems which can be outrun. The author suggests that a way to reduce the loss would be the use of coils into a ring, and the power supply to be made with many sources. A more precise study could identify other solutions.

To verify the results obtained theoretically, the author builds a two-phase microengine from ferrite, powered at 15 kHz, with a small tank filled with seawater as armature. At start-up, a piece with blades placed in the tank rotated with a few degrees in direction of swirling field, driven by the mobile liquid particles, and returning to equilibrium point at turning off the power. Similar experiments conducted with classical engines powered at 50 Hz didn't produce any results. The motion detected had very small amplitude, which can be explained with the very small powers and dimensions that were used. The experiment described above, although had a qualitative character, proves that the increase of frequency leads to the increase of the efficiency of an induction machine with liquid moving drive and so, the naval MHD induction thruster is doable.

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THE INFLUENCE OF THE SHALLOW WATERS TO SHIP'S HULL VERTICAL VIBRATION

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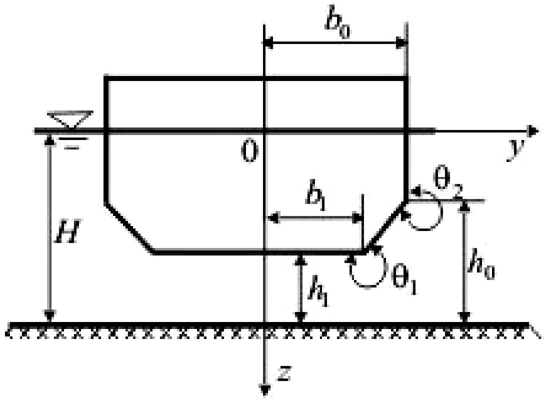
Abstract: *Hydrodynamic influence on ship's hull vibration in water is expressed in terms of added mass, which depends on flow domain, such as water depth, flow-field boundaries, ship's underwater shape and vibration modes. The problem of the added mass of a continuous ship-hull girder vibrating vertically in deep water and close to water bottom is studied. The added masses for vertical vibrations of the ship's hull are computed with the Schwarz-Christoffel method for shallow waters beneath the keel. The results obtained show differences between this method and the Lewis one, and also the added mass influence on the natural frequencies of a bulk carrier seen as a continuous girder. In order to determine the dynamic response of the ship's hull from the pulsatory pressures on the aft counter we utilized the dynamic transfer matrices method corresponding to the forced vibrations. Multiple personal computation programs has been realized for the computation of the necessary data (added mass through both methods, frequencies and vibration modes, responses to the forced vibrations, etc.). Numerical values show that the effects of shallow water are significant and quite important in ship's hull vibrations.*

Key words: *shallow water, ship vibration, added mass, transmission dynamic matrices method*

1 Schwarz-Christoffel method for computation of the added masses of ship's hull vertical vibrations in shallow waters

In the study of ship's hull vibratory motions, the presence of the fluid medium causes the increase of the ship's mass with the added mass value of the fluid which moved together with the vessel. This depends on the flow around the ship's hull, the ship's shape, the free surface, the vibration modes and the depth beneath keel. In present days the computation of ship's hull vibrations it's realized based on the added mass determined at infinite depth, with the Lewis method [1] help which was for the first time applied in 1927.

This section [1] presents the computation of the added masses of ship's hull vertical vibrations in shallow waters, using the Schwarz-Christoffel transform [9] for determination of the cross section of the hull in the bilge area and the Newman method [2].



Picture 1 Cross section geometrical parameters

- the intermediary domain, D_2, between $y \in (b_1, b_0)$ and $z \in [H - h_0, H - h_1]$;
- the external domain, D_3, between $y \in (b_0, \infty)$ and $z \in [0, H - h_0]$.

In the vertical vibrations with high amplitudes of the ship's hull, in which the speed is described by:

$$v = V\cos\omega t. \tag{1}$$

Assuming that the flow is irotational, the fluid's potential is given by:

$$\Phi(x, y, z) = \varphi(y, z)\cos\omega t \tag{2}$$

in which ω represents the angular frequency of the vertical vibrations, and V represents the speed's amplitude of the ship's vertical motions. In any point (*y, z*) of the studied fluid domain, the speed potential must concur with the next conditions:

1. $\nabla^2\varphi = 0$;
2. $\dfrac{\partial\varphi}{\partial z} = 0$ for *z=H*;
3. $\dfrac{\partial\varphi}{\partial z} - \upsilon\varphi = 0$, where $\upsilon = \dfrac{\omega^2}{g}$ at the free surface *(z=0)*;
4. $\dfrac{\partial\varphi}{\partial n} = V_n$, on the perpendicular direction on ship's hull;
5. $\lim\limits_{y\to\infty} \mathrm{Re}\left(\dfrac{\partial\varphi}{\partial y} - i\upsilon\varphi\right) = 0$.

The speed potential in the D_1 domain, in accord with the first *4* conditions and for $h_1 \to 0$, will have the next expression:

$$\varphi_1(y) = -\frac{V}{2h_1}y^2 + A_0 \tag{3}$$

in which can be noticed that in vertical direction the flow can be neglected.

To determine the speed potential in the intermediary domain the form of the transversal section in the ship's hull in the bilge area must be established by the θ_1 and θ_2 angles (picture 1):

$$\theta_1 = \pi + \frac{\pi}{n};\ \theta_2 = \frac{3\pi}{2} - \frac{\pi}{n} \tag{4}$$

in which n is arbitrary, and also transform the points from the real Z plane in source points from ζ plane, using the Schwarz-Christoffel transform:

$$Z = b_1 - \frac{ih}{\pi\beta^{\frac{1}{2}-\frac{1}{n}}} \int_1^{\zeta} \frac{(\zeta-1)^{\frac{1}{n}}(\zeta-\beta)^{\frac{1}{2}-\frac{1}{n}}}{\zeta} d\zeta + i(h_1 - H) \tag{5}$$

In relation (5) β is the point in ζ plane which corresponds to the intersection point between the bilge and the side plating from the Z plane, and was determined with:

$$\pi\beta^{\frac{1}{2}-\frac{1}{n}} t_0 = h_1 \int_1^{\beta} \frac{(\zeta-1)^{\frac{1}{n}}(\beta-\zeta)^{\frac{1}{2}-\frac{1}{n}}}{\zeta} d\zeta \tag{6}$$

in which t_0 means the length of the bilge (between $(b_1, H-h_1)$ and $(b_0, H-h_0)$):

$$t_0 = \sqrt{(b_0 - b_1)^2 + (h_0 - h_1)^2} \tag{7}$$

The complex potential in ζ plane is:

$$W(\zeta) = \frac{Q}{2\pi}\ln\zeta + C \tag{8}$$

in which Q represents the source intensity, and C is a constant that can be determined from the boundary conditions. For $\zeta \to 0$ relation (5) becomes:

$$Z = b_1 + \frac{h_1}{\pi}(\ln\zeta + K) \tag{9}$$

with the speed potential given by:

$$\varphi_2^i = \mathrm{Re}[W(\zeta)]_{\zeta\to 0} = \frac{Q}{2h_1}\left(y - b_1 - \frac{h_1}{\pi}K\right) + C \tag{10}$$

and for $\zeta \to \infty$ relation (5) becomes:

$$Z = -\frac{2ih_1\zeta^{\frac{1}{2}}}{\pi\beta^{\frac{1}{2}-\frac{1}{n}}} \tag{11}$$

with the speed potential given by:

$$\varphi_2^e = \mathrm{Re}[W(\zeta)]_{\zeta\to\infty} = \frac{Q}{\pi}\ln\frac{\pi r\beta^{\frac{1}{2}-\frac{1}{n}}}{2h_1} + C \tag{12}$$

where: $r = \sqrt{(y - b_0)^2 + (z - H)^2}$.

In the relations (9) and (10) K represents the integration constant and has the expression:

$$K = \left(\frac{1}{2} - \frac{1}{n}\right)\cdot\frac{1}{\beta} + \frac{1}{n} \tag{13}$$

In D_3 domain, the vertical vibration of the ship doesn't directly influence the fluid flow, but indirectly, caused by the movement of fluid masses between D_2 and D_3

domains, in both ways, movement which can be described as a q source in (b_0, H) point. The speed potential in D_3 domain, obtained by Wehausen and Laitone, accordingly with the stated conditions, is simplified neglecting the free surface effect. In these conditions the speed potential in D_3 domain, in the vicinity of D_2 is:

$$\varphi_3^i = \frac{q}{\pi}\left[\ln\left(\frac{\pi r}{H}\right) - 2\ln 2\right] \tag{14}$$

Constants A_0, q, Q and C are obtained from the continuity conditions between the domains:

$$\varphi_3^i = \varphi_2^e ; \varphi_2^i = \varphi_1^e ; \frac{\partial \varphi_2^1}{\partial y} = \frac{\partial \varphi_1^e}{\partial y} \tag{15}$$

The added mass given on a unit length, at the vertical vibration of the ship in the considered transversal section, is computed based on linear Bernoulli equation:

$$\overline{m_a} = \frac{2\rho}{V}\left[\int_0^{b_1} \varphi_1 dy + \int_{b_1}^{b_0} \frac{\varphi_2 (b_0 - b_1)}{t_2} dy\right] \tag{16}$$

from which:

$$\overline{m_a} = \overline{Cm_a} \cdot \frac{\rho \pi b^2}{2} \tag{17}$$

in which the added mass coefficient is:

$$\overline{Cm_a} = \frac{4b_1^2}{\pi b^2}\left[\frac{b_1}{3h_1} + \frac{K}{\pi} - \frac{2}{\pi}\left(1 + \frac{t_0}{b_1}\right)\ln\frac{h_1}{2(T + h_1)^{\frac{1}{2} - \frac{1}{n}}}\right] \tag{18}$$

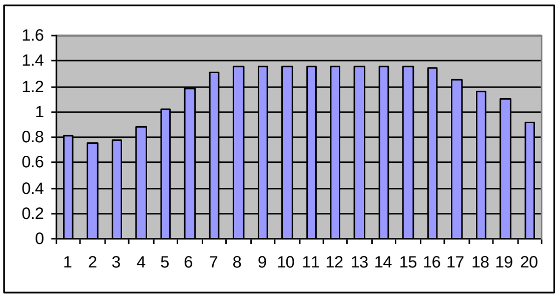
The terms b_{1i} and t_{0i}, for $i = \overline{1,20}$, have been determined from the ship's lines drawing.

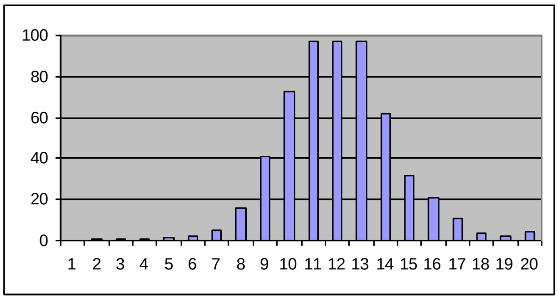
For the numerical computations a bulk carrier has been utilized (length L_{CWL}=296.33*m*, breadth B=46*m*, draft d=18*m*) divided in 20 segments of different lengths, with constant geometrical and mechanical characteristics [4].

The added masses have been determined, neglecting the 3*D* effect, with the classical Lewis method for infinite water depth and with the method presented above.

At the determination of expression (18), with $\frac{h_1}{b_{1i}} \leq 0.3...0.4$, in order to respect the imposed conditions, for the *U* shape transversal sections 5 significant depths beneath the keel: 0.09*m*, 0.2*m*, 0.5*m*, 0.9*m*, 1.8*m* have been chosen.

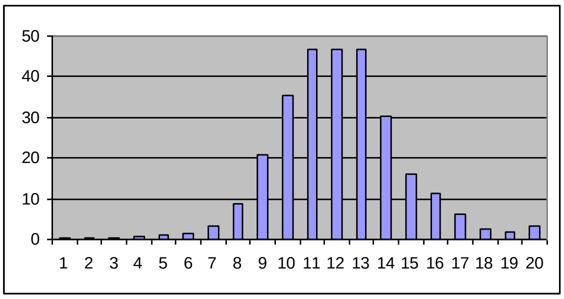
Using the Lewis method the *MASADL* program [4], has been realized, the obtained results being showed in the picture 2., and for the method presented the *MASADN* program [4, has been realized]. Pictures 2÷7 represent the variations of the added mass coefficients in the 20 segments of the ship, accordingly with the different depths beneath the keel.

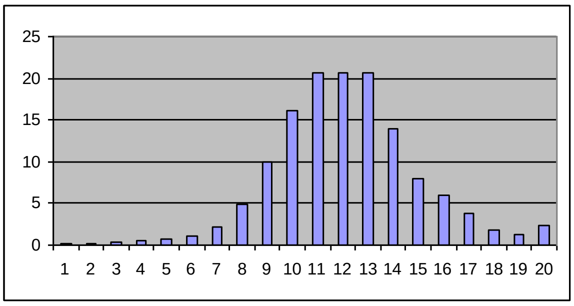




Picture 2 $\overline{Cm_a}$ *Lewis*

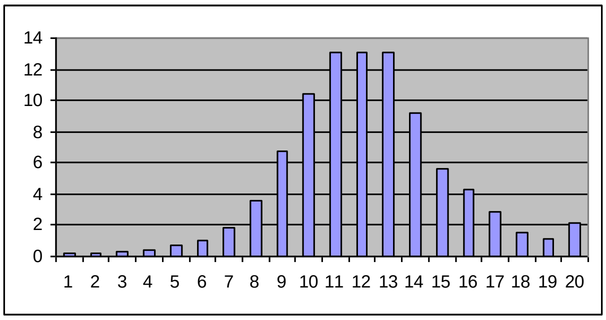
Picture 3 $\overline{Cm_a}$ *Newman* $h_1 / d = 0.005$

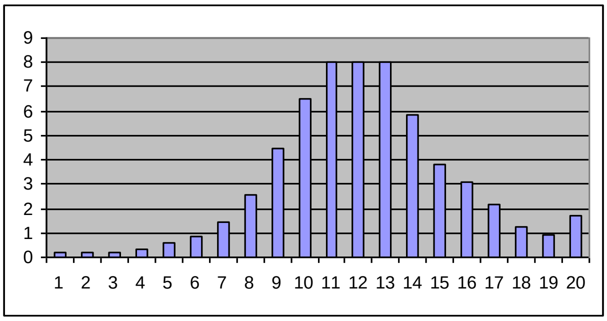




Picture 4 $\overline{Cm_a}$ *Newman* $h_1 / d = 0.011$

Picture 5 $\overline{Cm_a}$ *Newman* $h_1 / d = 0.028$





Picture 6 $\overline{Cm_a}$ *Newman* $h_1 / d = 0.05$

Picture 7 $\overline{Cm_a}$ *Newman* $h_1 / d = 0.1$

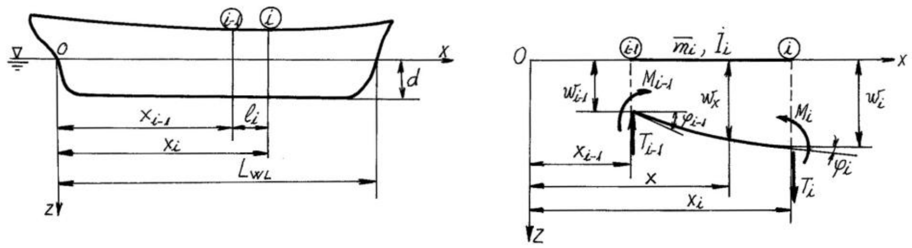
Drawing a parallel between the above results, it can be easily observed that the $\overline{Cm_a}$ values are far bigger in shallow waters (especially amidships), and also a modification in the $\overline{Cm_a}$ longitudinal distribution across the ship's length.

2 Natural frequencies of the ship-hull girder vibrating vertically in shallow waters

The natural frequencies of the continuous girder which represents the 170000tdw bulk carrier, has been computed using a transmission dynamic matrices method [5], [7] for the vertical vibrations, in the next hypothesis:

- the ship's hull is considered as a continuous girder of variable transversal section, free at the extremities, leaned on an elastic medium;
- each segment of the girder is modeled using the Euler girder theory, without being taken in consideration the rotational inertia and the shearing deformations;
- the vertical vibration is considered irrespective of the other general vibration types of the ship's hull.

It is considered a segment of the ship between the transversal sections i-1, respectively i (picture 8.a), design as a crossbeam with constant section, that bends in vertical plane (picture 8.b). The next notations are made: $\overline{m}_i = \overline{m}_{ni} + \overline{m}_{ai}$ represents the ship's mass and the added mass per length unit; I_i – the inertial momentum of the transversal section resistant to the longitudinal-vertical bending of the ship, computed to the neutral axis; w_{i-1}, w_i – the linear deflection in sections *i-1, i*; φ_{i-1}, φ_i – the angular deflection in sections *i-1, i*; M_{i-1}, M_i – the bending moments in sections *i-1, i;* $T_{z_{i-1}}$, T_{z_i} - the shearing forces in sections *i-1, i.*



Picture 8 The ship sections and longitudinal-vertical bending

The parameters of the two sections can be arranged in the next column vectors:

$$\left\{\overline{z}_{i-1}\right\} = \left\{w_{i-1}, \varphi_{i-1}, M_{i-1}, T_{z_{i-1}}\right\}^T; \quad \left\{\overline{z}_{i}\right\} = \left\{w_{i}, \varphi_{i}, M_{i}, T_{z_{i}}\right\}^T \tag{19}$$

Based on the differential equation of the free vertical vibrations, which takes into considerations only the bending deformations and the translation inertia of the masses [5][7]:

$$EI\frac{\partial^4 w(x,t)}{\partial x^4} + \overline{m}_i \frac{\partial^2 w(x,t)}{\partial t^2} = 0 \tag{20}$$

The solution of the equation is:

$$w(x,t) = w_x \cos(pt - \theta) \tag{21}$$

where the amplitude w_x, using Krilov fuctions, is

$$w_x = AS_{a_i x} + BT_{a_i x} + CU_{a_i x} + DV_{a_i x}. \tag{22}$$

Non-dimensional parameters from the section *i* are computed with:

$$[\overline{z_i}] = [\overline{A_i}] \cdot [\overline{z_{i-1}}] \tag{23}$$

in which $[\overline{A_i}]$ represents the non-dimensional dynamic transfer matrix.

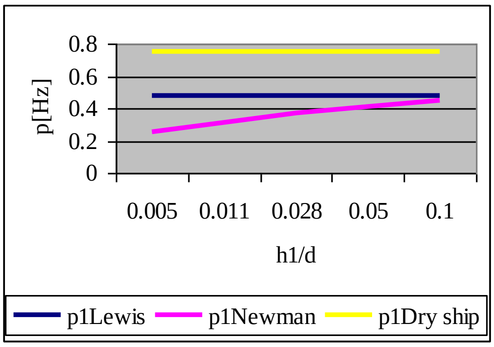
After the non-dimensional dynamic transfer matrix has been formed for all the *n* sections of the ship, the non-dimensional parameters from the section *n* (bow), function of the non-dimensional parameters from the origin (stern).

The engine values of the hull – ship vertical vibrations are calculated with:

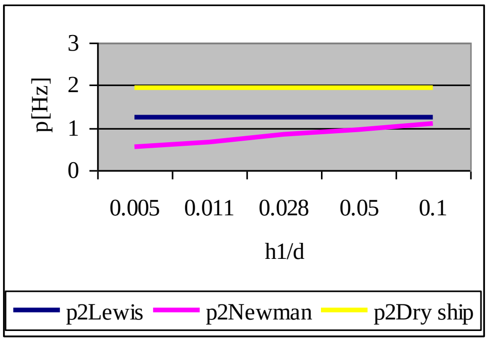
$$p = \left(\frac{\overline{a}}{l_0}\right)^2 \sqrt{\frac{EI_0}{\overline{m}_0}}, \tag{24}$$

where $\overline{a}$ results from the characteristic equation.

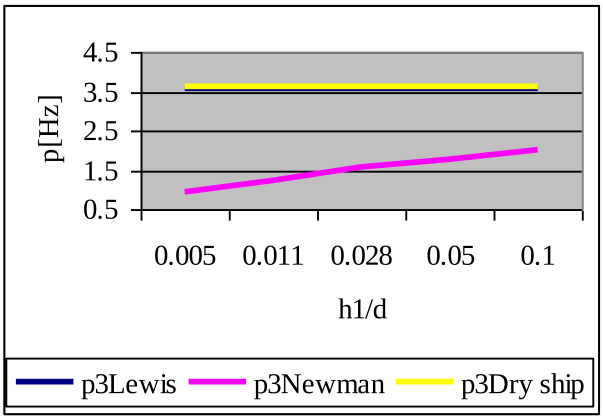
In the pictures 9, 10, 11 are represented by comparison the first three natural frequencies of the ship's hull, with and without the effect of added masses computed with the two methods presented in the second part of the article.



Picture 9 The natural frequencies in the first vibration way



Picture 10 The natural frequencies in the second vibration way



Picture 11 The natural frequencies in the third vibration way

From the representations can be easily seen that in the first vibration way (picture 9), computed for shallow waters, the frequency falls by maximum 50% from the frequency in infinite waters, also falls by approximately 70% from the frequency of the ship's hull without the effect of the added mass. These decreases in frequency become more prominent with the increase of the vibration mode. In this way, in the third vibration way, the frequency computed for shallow waters decreases by 75% from the frequency in infinite waters, also by 85% from the frequency of the ship's hull without the effect of the added mass.

3 The dynamic response of the ship's hull in shallow water

In the un-stationary cavitation case, because of the fast variation of the cavern with the blade's spinning angle, the amplitudes of the pressure fluctuations are very high (larger appreciatively with one order of size then the pulsations pressure amplitudes generated by the uncavitational propeller), being inverse-proportional with the distance between the blade and the ship's hull.

In this way, the un-stationary cavitation of the propeller represents one of the most significant sources of the ship's hull excitation, reason for what it is imperative to determine the pressure pulsation amplitudes on the stern counter in the design stage.

Starting with the hydrodynamic model of the pressures induced on the aft counter by the cavitational propeller, the Classification Societies has adopted methods more ore less simplified for the calculus of these pressures: the shipwake diagram method, the RNR method, the DNV method, etc.

One of the operative methods to determine the pressure pulsations on the aft counter, at the blade frequency and the double of it, useful even for distances less then 1 meter, is the DNV method [4].

In order to determine the dynamic response of the ship's body at pulsatory pressure generated by the propeller on the aft counter, using dynamic transfer matrix method, has been considered 3 different loading situations of the propeller combined with 2 loading situations of the ship (with full cargo, respectively in ballast), represented in table 1, for which has been computed the vertical components of the surface forces resultant.

Table 1 The ship and propeller loading

Nr. crt	n[rot/min]	ω_p[rad/s]	F_s[kN]	
			Δ=200462 t d_A=18m	Δ=90615 t d_A=10,09m
1	70	29,321	159,5	194,74
2	80	33,51	208,4	254,45
3	91	38,118	269,6	329,17

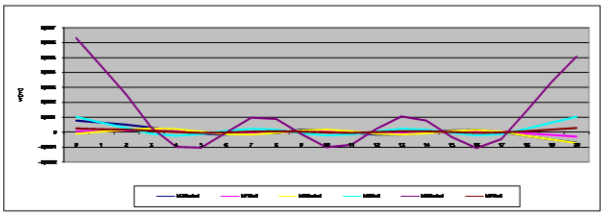
Taking in consideration the emplacement of the propeller at the stern, it is considered as application point of the surface forces the first node. In this way, in this node it is introduced a node matrix as:

$$[\overline{S}_1] = [0 \quad 0 \quad 0 \quad \overline{F_S}], \tag{25}$$

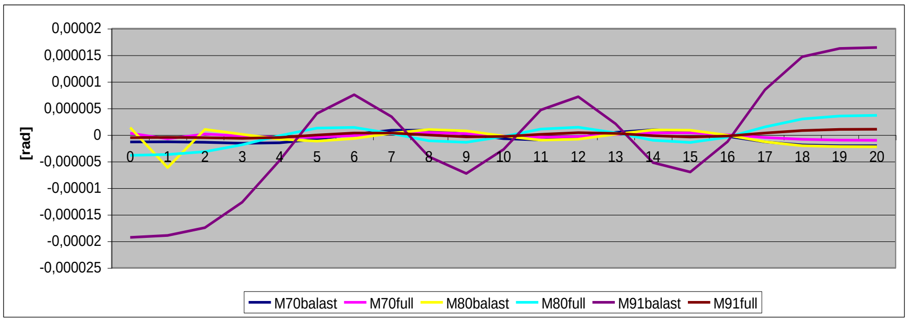
in which $\overline{F_S}$ represents the vertical non-dimensional component of surface forces resultant on the aft counter

$$\overline{F_S} = F_S \frac{l_0^2}{E_0 I_0}. \tag{26}$$

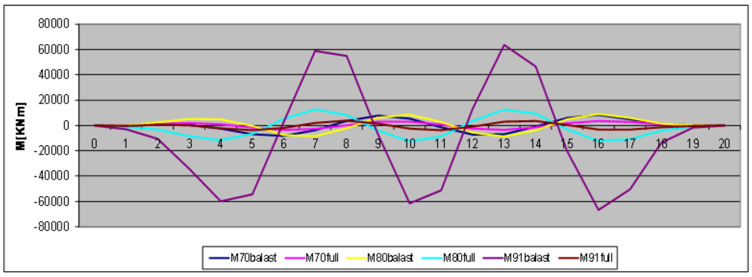
The results are shown in pictures 12, 13, 14 and 15.



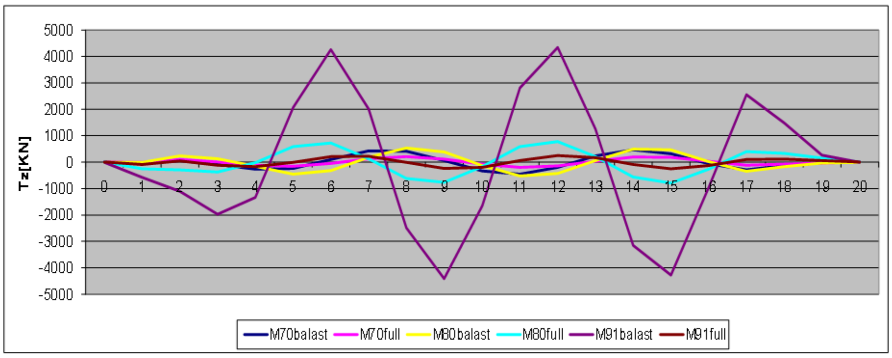
Picture 12 The displacement amplitudes



Picture 13 The spinning amplitudes



Picture 14 The bending moments amplitudes



Picture 15 The shearing forces amplitudes

From the presented graphics the high variations of the shifting amplitudes for different loading situations can be easily seen, respectively different depths beneath the keel considered.

4 Conclusions

For the numerical computations has been utilized a bulk carrier divided in 20 segments of different lengths, with constant geometrical and mechanical characteristics.

The added masses have been determined, neglecting the 3*D* effect, with the classical Lewis method for infinite water depth and with the Schwarz-Christoffel method.

Using the Lewis method the *MASADL* program, has been realized and for the Schwarz-Christoffel method presented in the second part of the article, has the *MASADN* program been realized.

For the *U* shape transversal sections 5 significant depths beneath the keel: 0.09*m*, 0.2*m*, 0.5*m*, 0.9*m*, 1.8*m* has been chosen.

The natural frequencies of the continuous girder which represent the 170000tdw bulk carrier, has been computed using the transmission dynamic matrices method for the vertical vibrations. We can lasing see, from the representation, that in the way, computed for shallow waters, the frequency falls by maximum 50% from the frequency in infinite waters, also falls by approximately 70% from the frequency of the ship's hull without the effect of the added mass. These decreases in frequency become more prominent with the increase of the vibration mode. In this way, in the third vibration way, the frequency computed for shallow waters decreases by 75% from the frequency in infinite waters, also by 85% from the frequency of the ship's hull without the effect of the added mass.

In order to determine the dynamic response of the ship's body at pulsatory pressure generated by the propeller on the aft counter, using dynamic transfer matrix method, has been considered 3 different loading situations of the propeller combined with 2 loading situations of the ship (with full cargo, respectively in ballast), represented in table 1, for which has been computed the vertical components of the surface forces resultant.

From the presented graphics can be seen the high variations of the shifting amplitudes for different loading situations, respectively different depths beneath the keel considered.

Numerical values show that the effects of shallow water are significant and quite important in ship's hull vibrations.

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A GENERALIZATION OF LEIBNIZ – NEWTON FORMULA

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Abstract. *The importance of definite integral in practice, in general, and for mathematics, particularly, it is well-known, therefore his calculus was a base problem of mathematics. In this paper we will make a generalization of calculus of the definite integral.*

Key words. *Integral, Leibniz – Newton*

Proposition 1. Let $f:[a,b]\to R$ a bounded function and $\sigma(f,P,(\xi_i))$ the associated Riemann sum of the function f, of the partition P of the interval $[a,b]$ and to the family of arbitrary points $(\xi_i)_{i=\overline{1,n}}$, then $\lim\limits_{n\to\infty}\sigma(f,P,(\xi_i))=\int_a^b f(x)\,dx$.

Lemma 1. $\lim\limits_{n\to\infty}\frac{b-a}{n}\sum\limits_{i=1}^{n} f\left(a+(b-a)\frac{i}{n}\right)=\int_a^b f(x)\,dx$

Proof. Let $P=(a=x_0<x_1<\ldots<x_n=b)$ an equidistant partition of the interval $[a,b]$. For $\xi_i\in[x_{i-1},x_i]$, we consider the particular form $\xi_i=x_i=a+\frac{b-a}{n}i$. Since the general form of the Riemann integral sum is $\sigma(f,P,(\xi_i))=\sum\limits_{i=1}^{n} f(\xi_i)(x_i-x_{i-1})$, then taking into consideration the considered particularities, we obtain the particular Riemann sum $\frac{b-a}{n}\sum\limits_{i=1}^{n} f\left(a+\frac{b-a}{n}i\right)$. Then, from Proposition 1, we have:

$$\lim_{n\to\infty}\frac{b-a}{n}\sum_{i=1}^{n} f\left(a+(b-a)\frac{i}{n}\right)=\int_a^b f(x)\,dx.$$

Proposition 2. (**Leibniz – Newton Formula**)

Let $f:[a,b]\to R$ such that

a) f is an integrable function on $[a,b]$;

b) f admit primitive on $[a,b]$.

If F is a primitive of f on $[a,b]$, then:

$$\int_a^b f(x)\,dx=F(b)-F(a).$$

Proof. Let $\sigma(f,P,(\xi_i))=\sum_{i=1}^{n} f(\xi_i)(x_i-x_{i-1})$ the Riemann sum associated to the function f, partition P of the interval $[a,b]$ and family of arbitrary points $(\xi_i)_{i=\overline{1,n}}$. Let $F:[a,b]\to R$ be a primitive of the function f. Since F is a primitive of the function f on $[a,b]$, f is continuous and derivable on $[a,b]$. Thus, f is continuous and derivable on any interval $[x_{i-1},x_i]$. On the interval $[x_{i-1},x_i]$, we can apply the theorem of Lagrange to function $F(x)$ and we obtain:

$$F(x_i)-F(x_{i-1})=f(\xi_i)(x_i-x_{i-1}),$$

hence:

$$\sum_{i=1}^{n}\left(F(x_i)-F(x_{i-1})\right)=\sum_{i=1}^{n} f(\xi_i)(x_i-x_{i-1}),$$

and:

$$\sum_{i=1}^{n} f(\xi_i)(x_i-x_{i-1})=F(b)-F(a).$$

Taking to limit in this relation and from Proposition 1, we have:

$$\int_a^b f(x)\,dx=F(b)-F(a).$$

Observation 1. To the bounded function $f:[a,b]\to R$ and to any partition P of the interval $[a,b]$ corresponds an infinity of continuum power of Riemann sums.

Lemma 2. Let $f:[a,b]\to R$ bounded and $M=\sup_{x\in[a,b]} f(x)$, $m=\inf_{x\in[a,b]} f(x)$. Then:

$$m(b-a)\le S(f,P)\le\sigma(f,P,(\xi_i))\le S(f,P)\le M(b-a).$$

Taking into account Lemma 2, the set $\{\sigma(f,P,(\xi_i))\mid(\xi_i)\in\xi_P\}$ is lower and upper bounded, where $(\xi_i)_{i=\overline{1,n}}$ is a family of arbitrary points, and ξ_P the set of all arbitrary points associated with the division P.

It is know that any continuous function is integrable, but conversely is not true.

Next, we evidentiate a class of functions which are not continuous on $[a,b]$, but they are integrable on $[a,b]$.

Definition 1. Let $A\subset R$ a set of real numbers. The set A is a negligible set (null Lebesgue measure set), if $\forall\ \varepsilon>0$, exist a sequence of intervals $(I_n)_{n\ge1}$ such that

a) $\bigcup_{n=1}^{\infty} I_n\supset A$; b) $\sum_{n=1}^{\infty}\|I_n\|<\varepsilon$ ($\|I_n\|$ the length of the interval I_n).

Definition 2. Function $f:[a,b]\to R$ is continuous almost everywhere on $[a,b]$, if the set of points A_0 from $[a,b]$ where f is not negligible.

Proposition 3.

a) A - negligible set $\Rightarrow \overset{0}{A}$ - negligible set.
b) A - set with a finite number of points is negligible.
c) If A is a countable set, then A is a negligible set.
d) Any finite or countable reunion of negligible sets is negligible.

Using negligible set and its properties we can prove the next Proposition.

Proposition 4. (Lebesgue Criterion) Let $f:[a,b]\to R$ be a bounded function. Function f is integrable on $[a,b]$ if and only if is continuous almost everywhere on $[a,b]$.
Taking into account this criterion, we study the problem if for such a functions,
Can $\int_a^b f(x)\,dx$ be computed with the help of the Leibniz – Newton formula?
Another frequently met problem is the following.
There are functions $f:[a,b]\to R$ which admits primitive on $[a,b]$, but practically these primitives can not be determinates on $[a,b]$, but only on intervals of the form $[a,c]\subset[a,b]$ or $[c,b]\subset[a,b]$. For such a functions there can be used Leibniz – Newton formula to compute $\int_a^b f(x)\,dx$?
The answer to these questions is given by the following Proposition.

Proposition 5. (Leibniz – Newton generalized formula)
Let $f:[a,b]\to R$ such that:

a) f is integrable on $[a,b]$.

b) f admits primitive on $[a,b]$.

c) Let $c\in(a,b)$ and the function $F:[a,b]\setminus\{c\}\to R$ derivable with the property $F'(x)=f(x)$, $\forall\ x\in[a,b]\setminus\{c\}$. Then:

$$\int_a^b f(x)\,dx=F(b)-F(a)+F(c-0)-F(c+0),$$

with $F(c-0)=\lim\limits_{x\nearrow c}F(x)$ and $F(c+0)=\lim\limits_{x\searrow c}F(x)$.

Proof. Let G a primitive of f on (a,b). Then $F'(x)=G'(x)=f(x)$, $\forall\ x\in(a,c)$. Hence, $F(x)=G(x)+c_1$, $\forall\ x\in(a,c)$. Similarly, $F(x)=G(x)+c_2$, $\forall\ x\in(c,b)$, $c_1\neq c_2$.

Since G is a primitive on (a,b), then G is continuous on (a,b), and G is continuous in $x=c$. Thus,

$$F(c-0)=G(c)+c_1 \text{ and } F(c+0)=G(c)+c_2.$$

Using Proposition 2, we have:

$$\int_a^b f(x)\,dx=G(b)-G(a).$$

Hence:

$$\int_a^b f(x)\,dx=F(b)-F(a)+c_1-c_2.$$

But:

$$c_1-c_2=F(c-0)+F(c+0).$$

Thus:

$$\int_a^b f(x)\,dx=F(b)-F(a)+F(c-0)-F(c+0).$$

Lemma 3. Let $f:[a,b]\to R$ bounded and discontinuous in $c_1<c_2<\ldots<c_p$, where $c_i\in(a,b)$, $i=\overline{1,p}$.

Then there exists $\int_a^b f(x)\,dx$ and $\int_a^b f(x)\,dx=F(b)-F(a)+\sum_{i=1}^{p}\left(F(c_i-0)+F(c_i+0)\right)$,

where $F'(x)=f(x)$, $\forall\ x\in[a,b]\setminus\{c_1,c_2,\ldots,c_p\}$.

Proof. Let $\alpha_i\in(c_i,c_{i+1})$, $i=\overline{1,p-1}$. We apply the formula from Proposition 5 on the intervals $[a,\alpha_1]$, $[\alpha_1,\alpha_2]$, …, $\left[\alpha_{p-1},b\right]$, and we add the results term by term and the proof is complete.

Example. Compute the integrals:

a) $\displaystyle\int_0^{2\pi}\frac{dx}{3+\cos x}$; b) $\displaystyle\int_0^2 f(x)\,dx$, where $f:[0,2]\to R$, $f(x)=\begin{cases}1, & x\in[0,2]\setminus\{1\}\\ -1, & x=1.\end{cases}$

Proof.

a) $f(x)=\dfrac{1}{3+\cos x}$ is continuous on R, thus admit primitive on R. Then $f(x)$ admit primitive on $[0,2\pi]$. But,

$$\int \frac{dx}{3+\cos x} = \frac{1}{\sqrt{2}} arctg \frac{tg\frac{x}{2}}{\sqrt{2}}, \quad \forall \; x \in (0,\pi) \cup (\pi, 2\pi)$$

because of the fact that $tg\frac{x}{2} = t$ is the only substitution which compute this primitive , is not defined in $x = \pi$.

Hence, $F(x) = \frac{1}{\sqrt{2}} arctg \frac{tg\frac{x}{2}}{\sqrt{2}}$ and $F(\pi - 0) = \frac{\pi}{2\sqrt{2}}$, $F(\pi + 0) = -\frac{\pi}{2\sqrt{2}}$. Then, according to Proposition 5:

$$\int_0^{2\pi} \frac{dx}{3+\cos x} = F(2\pi) - F(0) + F(\pi - 0) - F(\pi + 0) = 0 - 0 + \frac{\pi}{2\sqrt{2}} + \frac{\pi}{2\sqrt{2}} = \frac{\pi}{\sqrt{2}}.$$

b) By the virtue of Proposition 4, the function $f(x)$ is integrable, but does not admit primitive. But, we observe that $F : [0,2] \setminus \{1\} \to R$, $F(x) = x$ has the property $F'(x) = f(x)$, $\forall \; x \in [0,2] \setminus \{1\}$. Then, from Proposition 5, we have

$$\int_0^2 f(x)\,dx = F(2) - F(0) + F(1-0) - F(1+0) = 2 - 0 + 1 - 1 = 2.$$

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COMPUTATIONS OF SOME TRIGONOMETRIC SUMS USING MATRIX CALCULUS

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Abstract. *In mathematical models of some phenomena and optimization problems, we arrive to some sums of the form* $\sum_{k=1}^{n} a^k \sin kx$ *or* $\sum_{k=1}^{n} a^k \cos kx$. *In this paper, we will expose a method to compute this kind of sums using matrix calculus.*

Keywords: *matrix, trigonometric function*

Lemma. Let $a, b \in Q$ with $a^2 + b^2 < 1$ and $\exists p \in N$ such that $arctg\frac{b}{a} = \frac{\pi}{p}$ and

$$A = \begin{pmatrix} a & b \\ -b & a \end{pmatrix}$$

Then $A^p = cI_2$, $c \in R_-$.

Proof. It is obviously that $A = \sqrt{a^2 + b^2}\begin{pmatrix} \cos\frac{\pi}{p} & \sin\frac{\pi}{p} \\ -\sin\frac{\pi}{p} & \cos\frac{\pi}{p} \end{pmatrix}$ and $A^p = -\sqrt{(a^2 + b^2)^p} I_2$,

$c = -\sqrt{(a^2 + b^2)^p}$.

Problem. Compute the sums

$$\sum_{k=1}^{pn} b^k \cos k\varphi \text{ and } \sum_{k=1}^{pn} b^k \sin k\varphi .$$

Solution.

Let A a matrix such that

$$A^p = cI_2,\ A \in M_2(R) \tag{1}$$

and

$$A = b\begin{pmatrix} \cos\varphi & \sin\varphi \\ -\sin\varphi & \cos\varphi \end{pmatrix}.$$

We know that:

$$A^n = b^n \begin{pmatrix} \cos n\varphi & \sin n\varphi \\ -\sin n\varphi & \cos n\varphi \end{pmatrix}.$$

(2)

We consider the sum:

$$S = \sum_{k=1}^{pn} A^k = \begin{pmatrix} \sum_{k=1}^{pn} b^k \cos k\varphi & \sum_{k=1}^{pn} b^k \sin k\varphi \\ -\sum_{k=1}^{pn} b^k \sin k\varphi & \sum_{k=1}^{pn} b^k \cos k\varphi \end{pmatrix}.$$

(3)

In the same time, we have that:

$$A + A^2 + A^3 + \ldots + A^p = \begin{pmatrix} c_1 & c_2 \\ -c_2 & c_1 \end{pmatrix}.$$

(4)

On account of (1) and (4), we have

$$S = \left(A + A^2 + A^3 + \ldots + A^p\right) + A^p\left(A + A^2 + A^3 + \ldots + A^p\right) + A^{2p}\left(A + A^2 + A^3 + \ldots + A^p\right) +$$

$$+ \ldots + A^{(n-1)p}\left(A + A^2 + A^3 + \ldots + A^p\right) = \begin{pmatrix} c_1 & c_2 \\ -c_2 & c_1 \end{pmatrix}\left(1 + a + a^2 + \ldots + a^{n-1}\right) =$$

$$= \frac{1-a^n}{1-a}\begin{pmatrix} c_1 & c_2 \\ -c_2 & c_1 \end{pmatrix}.$$

(5)

Now, by (3) and (5), we have:

$$\sum_{k=1}^{pn} b^k \cos k\varphi = \frac{1-a^n}{1-a} c_1 \text{ and } \sum_{k=1}^{pn} b^k \sin k\varphi = \frac{1-a^n}{1-a} c_2$$

(6)

Example:

$$\sum_{k=1}^{99} 2^k \cos k\frac{\pi}{3} = ? \text{ and } \sum_{k=1}^{99} 2^k \sin k\frac{\pi}{3} = ?$$

Solution. Let

$$A = \begin{pmatrix} 1 & \sqrt{3} \\ -\sqrt{3} & 1 \end{pmatrix} = 2\begin{pmatrix} \frac{1}{2} & \frac{\sqrt{3}}{2} \\ -\frac{\sqrt{3}}{2} & \frac{1}{2} \end{pmatrix}.$$

Thus:

$$A = 2\begin{pmatrix} \cos\frac{\pi}{3} & \sin\frac{\pi}{3} \\ -\sin\frac{\pi}{3} & \cos\frac{\pi}{3} \end{pmatrix}.$$

Then:

$$A^k = 2^k\begin{pmatrix} \cos\frac{k\pi}{3} & \sin\frac{k\pi}{3} \\ -\sin\frac{k\pi}{3} & \cos\frac{k\pi}{3} \end{pmatrix}.$$

Let the sum:

$$S = \sum_{k=1}^{99} A^k = \begin{pmatrix} \sum_{k=1}^{99} 2^k \cos\frac{k\pi}{3} & \sum_{k=1}^{99} 2^k \sin\frac{k\pi}{3} \\ -\sum_{k=1}^{99} 2^k \sin\frac{k\pi}{3} & \sum_{k=1}^{99} 2^k \cos\frac{k\pi}{3} \end{pmatrix}. \tag{7}$$

But:

$$A + A^2 + A^3 + \ldots + A^p = \begin{pmatrix} -9 & 3\sqrt{3} \\ -3\sqrt{3} & -9 \end{pmatrix} \text{ and } A^3 = -8I_2. \tag{8}$$

Now, if:

$$S = \left(A + A^2 + A^3\right) + A^3\left(A + A^2 + A^3\right) + \ldots + A^{3\cdot 32}\left(A + A^2 + A^3\right),$$

taking into account (8), we have:

$$S = \begin{pmatrix} -9 & 3\sqrt{3} \\ -3\sqrt{3} & -9 \end{pmatrix}\left(1 + (-8)^3 + 8^6 - 8^9 + \ldots - 8^{93} + 8^{96}\right) =$$

$$= \begin{pmatrix} -9 & 3\sqrt{3} \\ -3\sqrt{3} & -9 \end{pmatrix}\left(1 + \left((-8)^3\right)^1 + \left((-8)^3\right)^2 + \left((-8)^3\right)^3 + \ldots + \left((-8)^3\right)^{31} + \left((-8)^3\right)^{32}\right) =$$

$$= \frac{1 - \left((-8)^3\right)^{33}}{1 - (-8)^3}\begin{pmatrix} -9 & 3\sqrt{3} \\ -3\sqrt{3} & -9 \end{pmatrix} = \frac{1 - 8^{99}}{513}\begin{pmatrix} -9 & 3\sqrt{3} \\ -3\sqrt{3} & -9 \end{pmatrix}.$$

Hence:

$$S = \begin{pmatrix} -\frac{1+8^{99}}{57} & \frac{1+8^{99}}{\sqrt{3}\cdot 57} \\ -\frac{1+8^{99}}{\sqrt{3}\cdot 57} & -\frac{1+8^{99}}{57} \end{pmatrix} \tag{9}$$

Now, by (7) and (9), we have:

$$\sum_{k=1}^{99} 2^k \cos k\frac{\pi}{3} = -\frac{1+8^{99}}{57} \text{ and } \sum_{k=1}^{99} 2^k \sin k\frac{\pi}{3} = \frac{1+8^{99}}{\sqrt{3}\cdot 57}.$$

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AN EXPERIMENTAL METHOD AND DEVICE DEMONSTRATING QUANTIFIED ABSORPTION OF ENERGY BY ATOMS

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Abstract: *In the paper the authors present a method and an experimental device for didactic purposes which allows the demonstration of quantified absorption of energy by mercury and oxygen atoms through Franck-Hertz method.*

Keywords: *energy, atoms, absorption*

1 Theoretical considerations

The Franck-Hertz experiment, now considered fundamental for confirmation Bohr's theory, regarding the quantification of atoms energy, was conducted for the first time by Jacob Franck and Gustav Hertz between 1912-1914 and awarded with the Nobel Prize in 1925.

The experiments started before the apparition of Niels Bohr first paper in 1913, and the conclusions were contrary to the statements made by Bohr [3;4].

Only in 1917, Davis and Gucher showed the difference between excitation and ionization of the atom, thus explaining the results obtained by Franck and Hertz according to Bohr's theory [4].

Basically, the mercury atoms put in an enclosure as gas at low pressure, were excited with electron bombardment. The electrons were produced by an incandescent cathode or by a photocathode as in the experiment made by Townsend. The electrons are accelerated by an electric field generated in the space between cathode and grid, where an adjustable voltage is applied. Accelerated electrons circulate though a braking field made between the grid and a very close anode, generating an anodic current, which can be measured with a sensitive galvanometer or an electronic meter. One can notice that when the acceleration voltage is increased from 0 to 4,9 V the current intensity is increased, and the shape of variation curve of current intensity as a function of grid-cathode voltage, is quite similar to the one of a vacuum triode with small slope. When is reached the 4,9 V voltage, the current intensity is dropping. The explication is: under 4,9 V the electrons with energy smaller than 4,9 eV, despite numerous collisions with the mercury atoms, manage to go through the potential barrier between grid and anode.

When the acceleration voltage reaches 4,9 V, the electrons will have a 4,9 eV energy in the vicinity of the grid, which is the exact energy necessary for the mercury atoms to make the first quantic transition (first excitation). In this moment, the mercury atoms absorb the energy of the electrons that cannot go through the potential barrier between grid and anode, and the anodic current has a pronounced decay. The current doesn't annul because of the static dispersion of electrons energy around value 4,9 eV. For voltages over 4,9 V, the collisions are again elastic, so the anodic current increases. When the 2x4,9V = 9,8V voltage is reached, the electrons suffer a first energy loss through nonelastic collisions at the half of the space between anode-grid and a second one right in the vicinity of the grid, so a new decrease of anodic current appears.

It is clear that the results of Franck-Hertz experiment are experimental proves regarding the existence of discrete levels of energy in atoms. One can reveal even the emissions of ultraviolet radiations emitted by the mercury atoms that go back to the fundamental state.

The experiment is made with a tube from which the air was cleared out and replaced with mercury vapors obtained through mercury vaporization 120 ^{0}C – 150 ^{0}C.

The experiment was conducted with thyratrons too (GT-0,1/0,3 and 884) [4].

The special educational valences of this experiment impose the presentation of this experiment to high school students in terminal years when they study atom structure and to students who take a course of general physics. When the special tube is missing because it is found only in specialized laboratories of atom physics or plasma physics, one can adopt different solutions [6].

In this paper, authors propose a different device using barometric vacuuming with mercury column.

2 The experimental device

The experimental device consists of:

- active enclosure with barometric vacuuming
- electronic measuring device
- ower circuits

2.1 Active enclosure with barometric vacuuming
2.1.1 The execution of active enclosure
2.1.1.1 The cathode filament

The filament has a shape of a plane spiral obtained by bringing in the same plane a conic spiral of copper nickel Φ = 0,35. The length of the spiral conductor was approximately 100 mm, which was heated around 700 K when a voltage of 4V for a current of 2A was applied.

Then the filament was activated with barium oxide through repeatedly decomposition over its surface of a suspension of $BaCO_3$ in a CO_2 atmosphere. The emission current calculated after procedures in paper [1] should be 30 μA, but during the experiments, we got a smaller value.

2.1.1.2 The accelerating grid

It was made in the same way as the cathode filament by flattening a conic spiral of a conductor of Cu-Ag Φ = 0,3 mm.

Both the cathode-filament and the accelerating grid are mounted on conductors separated by a pearl of glass obtained by melting a 3 glass capillary tubes.

2.1.1.3 The vacuuming enclosure

The enclosure where the barometric vacuuming takes place is a tube of glass a diameter of 25 mm and a length of 70 mm, ending in the inferior region with an inlet for a flexible tube. The filament-grid unit, fitted with joint conductors, is introduced through the superior part of the tube, after which this one is melted with a propane-oxygen burner, taking special protection measures for the joint conductors through the glass and for the glass too (slow and uniform cooling to avoid the tension and the cracking).

2.1.1.4 The anode

The anode is a stainless disk floating on the free surface of the mercury. The free surface of the mercury cannot be the anode because during the vacuuming is covered with a layer of oxide. The active enclosure in final form in represented in fig.1.

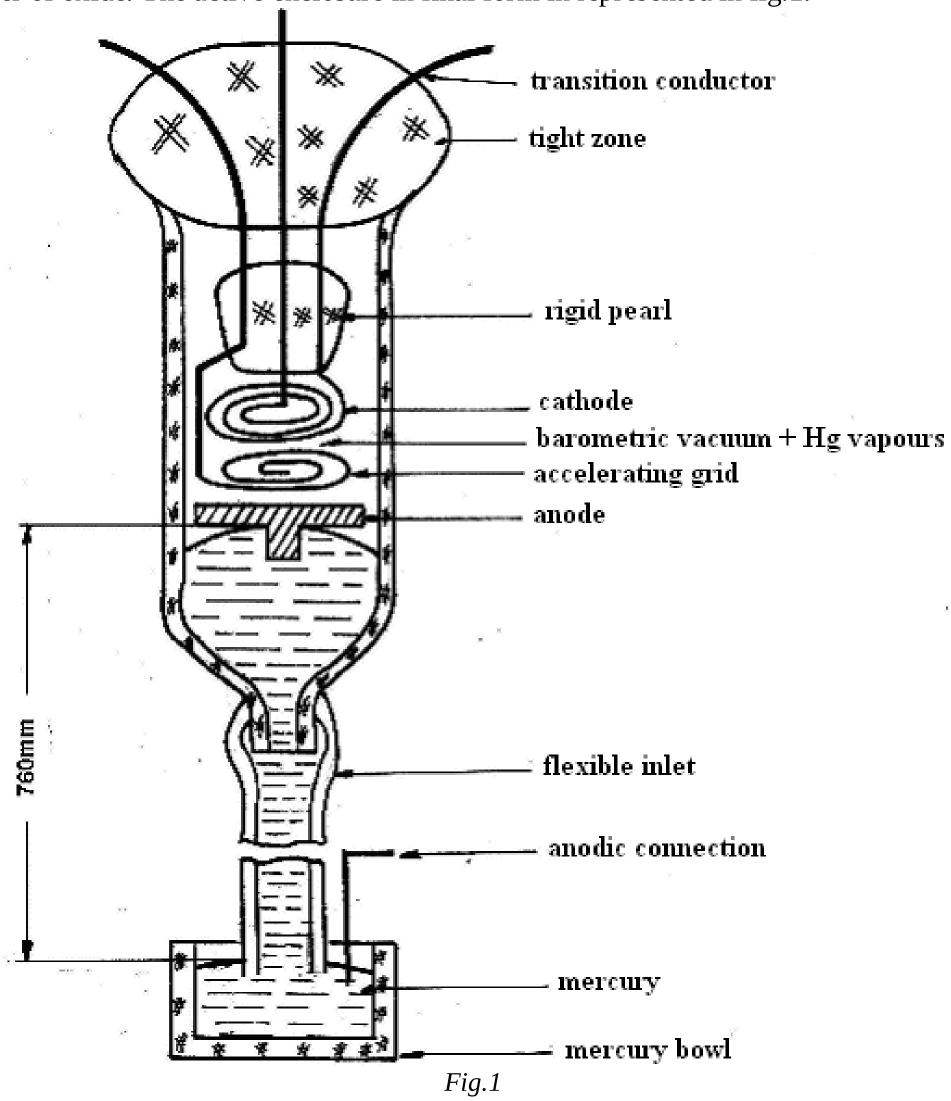

Fig.1

2.2 The electronic measuring device

To measure the intensity of the anodic current, the amplifier from fig.2 was used along with an electronic voltmeter IPT 01.

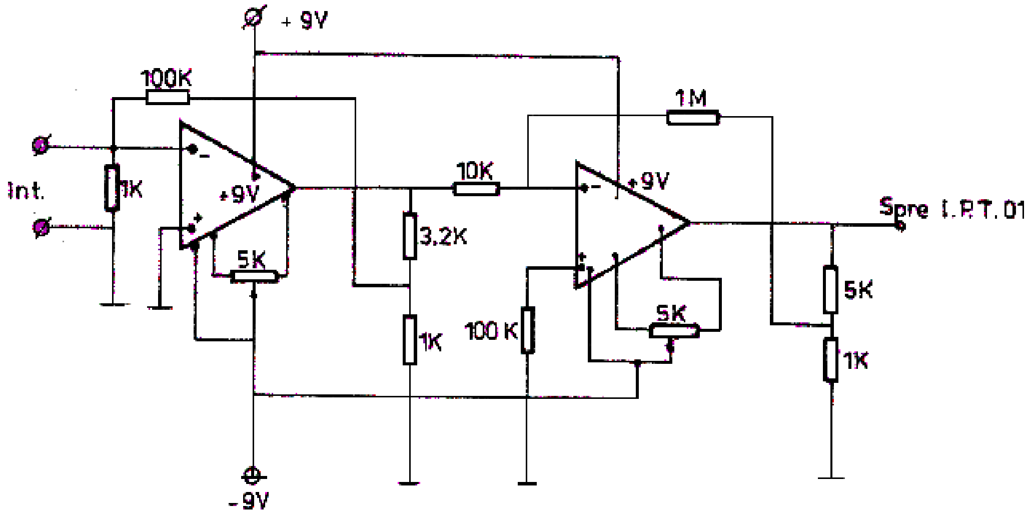

Fig.2

2.3 The power circuits

The acceleration voltage cathode-grid come from a rheostat powered with 3 batteries of 9V, and the braking voltage grid-anode come from a rheostat powered with 1 battery of 1,5V. Both circuits have voltmeters. The electric diagram is presented in fig.3.

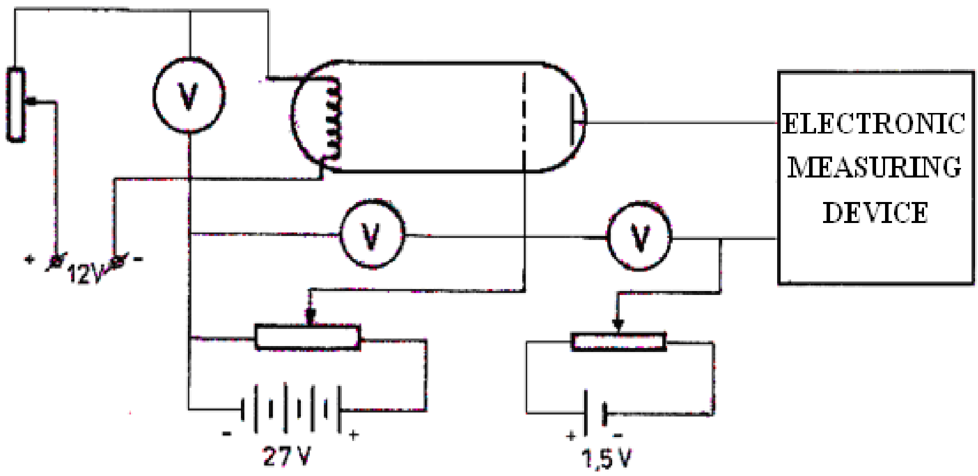

Fig.3

3. The experiment and the results
3.1 Tube vacuuming

3.1.1 Mercury is poured into the tube through the flexible inlet so the mercury will fill the whole tube.

3.1.2 The tube is left vertically with the filling hose upward to permit the evacuation of air bubbles adherent to balloon wall.

3.1.3 The end of the flexible inlet is introduced in a tank with mercury and the tube is rotated with 180^0 lifting to a height so it will form the barometric level in the active zone, leaving the anode to 0,5-1 mm from the grid.

3.2 Heating the filament

3.2.1 The intensity of the current of the filament is continuously increased so that the filament will heat up and the adherent mercury drops from the filament will disappear.

3.2.2 The filament stays heated to provoke the evaporation of the mercury and to increase the vapour pressure in the balloon.

3.3 Characteristics plotting

3.3.1 The existence of the electronic current produced by the cathode emission is verified by plotting the characteristic of the diode for the region filament-grid to be sure that the results are not affected by the residual current on the walls.

3.3.2 One realizes the circuit from fig.3, obtaining nonflat curves as in fig.4, where the braking voltage modifies between 0,5 V and 1,5 V.

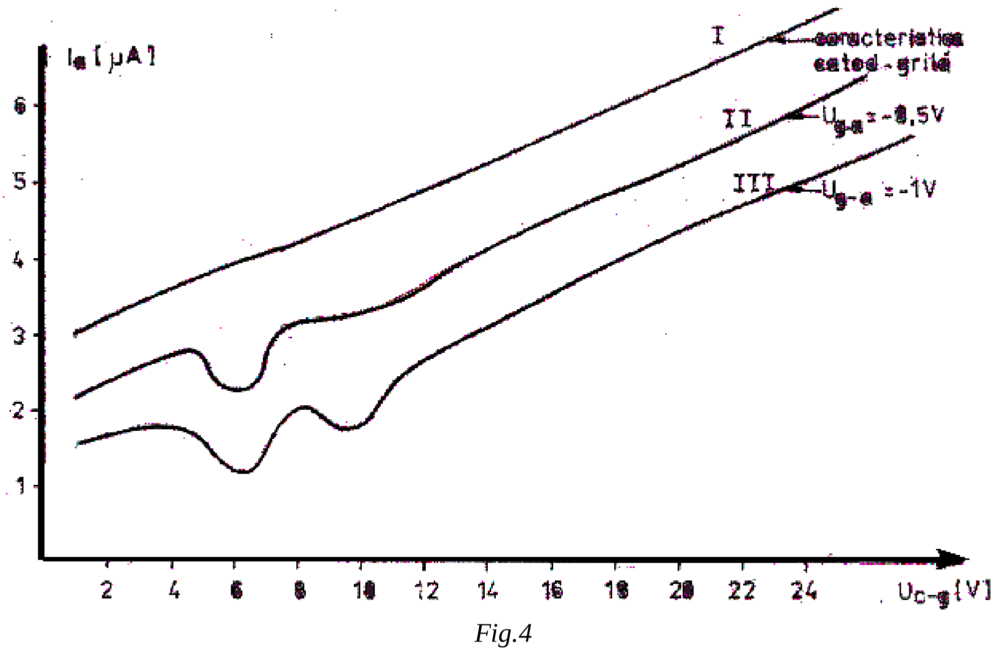

Fig.4

4 Conclusions

It is obvious the decrease of the current intensity in the region of 4,9V, around the expected value, which confirms the quantified absorption of the energy. The second minimum, expected around 9,8V, is not that obvious. For $U_{g-a} = -1V$, one can notice a second minimum, but it appears at a smaller value. The authors explain this by the fact that in the tube still remains a little air; plus, because of the decomposition of the mercury oxide from the filament oxygen appears. For curve II, oxygen has the first excitation potential and slope abatement at 7,9V; the obvious minimum of curve III in region 8-10V is a consequence of superposition the minimum of mercury, oxygen and probably of nitrogen.

In any case, the Franck-Hertz effect is obvious.

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THEORETICAL RESEARCHES OVER THE QUALITY FACTOR OF AN INDUCTION MHD THRUSTER

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Abstract: *Naval induction magneto hydrodynamic (MHD) thruster is an induction linear engine having as a moving drive the sea water. The efficiency of an induction engine can be characterized by its quality factor defined by E. Laithwaite as the ration between the magnetization reactance and the moving drive's resistance.*
In his paper, the author gives a different definition to this factor and demonstrates the equivalence between his definition and that of E. Laithwaite.
The author find out the mathematical expression for this factor in the case of a liquid moving drive whose speed is not constant on a perpendicular direction to the pipe.
In conclusion, the author compares the quality factor of an engine with sea water moving drive with that of an engine with a metallic moving dive and demonstrates the fact that the MHD thruster is feasible if the polar step and the frequency have increased sizes.

Keywords: *induction, MHD THRUSTER, research*

1 General considerations

Experiments conducted with linear induction machines with solid moving drive replaced by a pipe filled with seawater, with the frequency of 50 Hz and with the active power of 2 kW, showed a starting force of $2{,}89 \cdot 10^{-7}$ N. Forces of this magnitude are not capable to displace seawater or to move a small ship.[6] The quality factor of engine with seawater as indus is of 10^{-4} magnitude. In the case of classic induction engines, to obtain efficiency greater than 50 %, one must have a quality factor very over unitary.

So, it is needed a study about the quality factor for seawater moving drive, especially because in this case, the moving armature slip isn't constant.

2 A new definition of the quality factor of a linear induction machines

Because the naval MHD thruster is a linear induction machine, its efficiency is described using the quality factor Q, defined by Laithwaite as the ratio between magnetization reactance and moving drive's resistance:

$$Q = \frac{\omega L_2}{R_2} \tag{1}$$

From this definition results:

$$Q = \frac{2\tau_p \mu_0 f}{\pi \frac{\rho}{\Delta} h} \tag{2}$$

where: τ_p is the polar step, μ_0 is the permeability of vacuum, *f* is the supply frequency, ρ is resistivity, Δ is moving drive thickness and *h* is the size of air-gap. [1]

In case of an induction machine with seawater as moving drive, it is difficult to use eq. 1 because of obvious reasons; that's why we assume a redefinition of quality factor on energetic considerations.

The machine takes energy from the source and a part of it is dissipated in the liquid environment as active energy, and the other part is stocked as reactive energy accumulated in the magnetic field from air-gap. The author considers that the machine will be much efficient as the ratio between dissipated energy and accumulated energy will be greater, and he proposes to characterize the efficiency of naval MHD thruster using the quantity „energetic ratio" R.[4] If the dissipated Joule energy is W_j and the magnetic field energy in the air-gap is W_m, this ratio is:

$$R = \frac{W_j}{W_m} \tag{3}$$

The author will demonstrate that the energetic ratio R and the quality factor Q represent the same physical reality. To do this, one will calculate the energies W_j and W_m. The polyphased coil of inductor will generate in the air-gap a magnetic field with the aspect of progressive wave with intensity given by: [2;3]

$$H(x,t) = H_m \sin(\omega t - \frac{\pi x}{\tau_p}) \tag{4}$$

This field travel with a synchronous velocity towards a fixed point:

$$v_s = 2\tau_p f \tag{5}$$

If indus moves with the velocity v, the relative velocity of magnetic field towards indus will be:

$$v_r = v_s - v_0 \tag{6}$$

The progressive field covers a distance equal with a polar step during t_τ:

$$t_\tau = \frac{\tau_p}{v_s - v} \tag{7}$$

During this time, in moving drive a Joule energy having this expression will be dissipated:

$$W_j = \int_0^{t_\tau} P_j(x,t)dt \tag{8}$$

The infinitesimal Joule power dissipated in an moving drive element of length dx will be:

$$dP_j = u^2(x,t)dg(x) \tag{9}$$

where u(x,t) is the voltage induced by the progressive field in indus element situated at x towards a fixed point placed at the extremity of a pole, at moment t, and g(x) is indus conductivity:

$$g = \frac{1}{R_2} \tag{10}$$

A stripe of moving drive with an infinitesimal length dx, and with a width equal with the width „a“ o moving drive f, will have the next conductance:

$$dg = \frac{1}{\rho \frac{a}{\Delta dx}} = \frac{dx}{\frac{\rho}{\Delta} a} \tag{11}$$

The ratio $\rho / \Delta = \rho_s$ is surface resistance [1, 2].

The value of induced voltage will be given by the law of electromagnetic induction:

$$u(x,t) = -\frac{\partial \Phi(x,t)}{\partial t} - Bav_r \tag{12}$$

In a first stage we disregard the first element due to pulsation, and in this case the energy dissipated through Joule effect in an moving drive segment with length τ_p will be:

$$W_j = \int_0^{t_\tau} \int_o^{\tau_p} \frac{\mu_o^2 H^2(x,t) a v_r^2}{2\rho_s} dx dt \tag{13}$$

Replacing in 13 the expression of field intensity and making the integrations, we get:

$$W_j = \frac{\mu_o^2 H_m^2 a \tau_p^2 v_r}{2\rho_s} \tag{14}$$

The magnetic energy is obtained integrating the energy density of the magnetic field on air-gap volume:

$$W_m = \iiint_{V_{\text{int}\,refier}} \frac{\mu_0 H^2(x,t)}{2} dx dy dz \tag{15}$$

or:

$$W_m = \frac{\mu_0 H_m^2}{2} \int_0^a dy \int_0^h dz \int_0^{\tau_p} \sin^2(\omega t - \frac{\pi x}{\tau_p}) dx \tag{16}$$

After integration:

$$W_m = \frac{\mu_0 H_m^2 a \tau_p h}{4} \tag{17}$$

The value of energetic ratio R becomes:

$$R = \frac{2\mu_0 \tau_p v_r}{\rho_s h} \tag{18}$$

Introducing the armature slip:

$$s = \frac{v_s - v}{v_s} = \frac{v_r}{v_s} \tag{19}$$

and replacing the synchronous velocity 5, we obtain:

$$R = 2\pi s \frac{2\mu_0 f \tau_p^2}{\pi \rho_s h} = 2\pi s \frac{2\mu_0 f \tau_p^2}{\pi \frac{\rho}{\Delta} h} \tag{20}$$

In the moment of start-up, s = 1, and we get the ratio:

$$\frac{R}{Q} = 2\pi \qquad (21)$$

(in fact, relation 2 was deducted in the moment of start-up too).

The fact that R/Q is a non-dimensional constant, it allows us to assert that both R and Q express the same physical reality: the efficiency of linear induction machine.

The fact that R≠Q can be a coincidence of the fact that pulse induced voltage wasn't taken into account. So, quantity R can be used instead of Q, with the advantage of a clearer physical mean and a easier calculus.

3. The quality factor of the linear induction machine using sea water as naval thruster

In the case of the naval MHD thruster, the phenomena are getting complicated because the velocity in the flow channel is not uniform, and it has a maximum value on the channel axis and it is zero at contact with walls. To apply eq. (9.3) in the case of a thruster with seawater, the energies W_j and W_m must be calculated using eq. Maxwell – Hertz with respect to hydrodynamic processes expressed with eq. Navier - Stokes. [2] The resulting differential equations systems are very complex and can be solved only using simplifying hypothesis. Such calculus is described by the author in paper [4], but the results couldn't be verified experimentally. Because of this, he proposes another simpler calculus method which leads to similar conclusions. The armature slip *s* can be defined only locally, so it will be a function like this:

$$s(x,z) = \frac{v_s - v(x,z)}{v_s} = 1 - \frac{v(x,z)}{v_s} \qquad (22)$$

In the proximity of free surface and on the axis of MHD channel, velocity has a maximum value $v_{o\max}$ and the moving armature slip is:

$$s_0 = \frac{v_s - v_{0\max}}{v_s} = 1 - \frac{v_{0\max}}{v_s} \qquad (23)$$

The energy ratio R can be expressed using the mean value of armature slip:

$$\bar{s} = \frac{1}{2ah}\int_{-a}^{+a}\int_{0}^{h} s(x,z)dxdz \qquad (24)$$

Velocity value in the channel, depending on maximum velocity $v_{0\max}$, is calculated in [4]:

$$v(x,z) = v_{0\max}\frac{1-\dfrac{chHx}{chHa}}{1-\dfrac{1}{chHa}}\frac{z^2}{h^2} \qquad (25)$$

Ha is Hartman's number:

B_0 is here the actual value of magnetic field induction:

$$B_0 = \frac{B_m}{\sqrt{2}} = \frac{\mu_0 H_m}{\sqrt{2}} \qquad (27)$$

$\sigma = \dfrac{1}{\rho}$ is the environment conductivity (for seawater $\tau \cong 5\Omega^{-1}m^{-1}$) and η is the dynamic viscosity. The factor z^2/h^2 in 25 appeared because one assumed the simplifying hypothesis that the depth distribution of velocities is very little influenced by the magnetic field.

Replacing in 24 the value of armature slip from 25, one gets:

$$\bar{s} = \frac{1}{3}\left[2 + s_0\left(1 - \frac{thHa}{Ha}\right)\right] \tag{28}$$

Replacing 28 in the expression of energy ratio 20, one gets:

$$R = \frac{2\pi}{3}\frac{2\mu_0 f\tau_p^2}{\pi\frac{\rho}{\Delta}h}\left[2 + s_0\left(1 - \frac{thHa}{Ha}\right)\right] \tag{29}$$

Of course, this equation is not entirely correct because it was deduced using simplifying hypothesis, but it offers some information regarding the phenomena produced by the irregularity of flowing velocity in the channel.

A first resulting information is that when the flowing velocity on the channel axis becomes zero, then the moving armature slips is zero; but R isn't zero like in machine with solid moving drive. In this case, although the inductive processes disappear on channel axis, they remain present elsewhere. The generating forces will differ from zero, causing water drive. The value of energy ratio R is mainly determined by the element in front of eq. 29 which increases quadratic with polar step and linear with frequency. The main drawback of the naval MHD thruster comes from the high value of seawater resistivity. For example:

$$\frac{\rho_{seawater}}{\rho_{copper}} \cong 10^7 \tag{30}$$

To have the same magnitude degree for energy ratio R in MHD propulsor like in an engine with copper indus, it is necessary to satisfy the eq.:

$$\frac{(\tau_p^2 f)_{seawater}}{(\tau_p^2 f)_{copper}} \cong 10^7 \tag{31}$$

Eq. 31 explains the fact that the experiments made with inductors designed after classical principals in which the metallic indus was replaced with a pipe with seawater, didn't lead to concluding results.

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SOLVING A GAUSS – KUZMIN THEOREM FOR RCF USING RSCC

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Abstract: *The paper present a solution for a Gauss – Kuzmin theorem associated to the regular continued fractions using the ergodic behavior a homogeneous random system with complete connections associated with this expansion.*

Key words: *continued fractions, random system with complete connections , Gauss – Kuzmin*

1 Introduction

Let Ω denote the collection of irrational numbers in the unit interval $I = \ 0,1 \ $. Write $x \in \Omega$ as a regular continued fraction (RCF):

$$x = \cfrac{1}{a_1\ x\ + \cfrac{1}{a_2\ x\ + \ddots}} := \left[a_1\ x\ , a_2\ x\ , \ldots \right] \tag{1.1}$$

where the $\mathbf{N}_+$- valued functions on Ω, $a_k\ x\ $, are unique and called incomplete quotients of $x \in \Omega$. We will usually drop the dependence on x in the notation of incomplete quotients a_n, $n \in \mathbf{N}_+ := \ 1,2,\ldots\ $.

Define on I the transformation τ as follows:

$$\tau\ x\ := \tau\ \ a_1, a_2, \ldots\ \ = \ a_2, a_3, \ldots\ ,\ x \neq 0;\ \tau\ 0\ := 0 \tag{1.2}$$

It follows from (1.1) and (1.2) that for $x \neq 0$ we have:

$$x = \frac{1}{a_1 + \tau\ x} \tag{1.3}$$

Consequently, we can write the transformation τ of $\ 0,1\ $ as:

$$\tau\ x\ := \frac{1}{x} - \left\lfloor \frac{1}{x} \right\rfloor,\ x \neq 0 \tag{1.4}$$

where $\lfloor \cdot \rfloor$ denotes the floor function. Usually, this transformation is called the *Gauss map*. For $x \neq 0$, we get:

$$a_1 = \left\lfloor \frac{1}{x} \right\rfloor \text{ and } a_n = a_1\ \ \tau^{n-1}\ x\ \ ,\ n \in \mathbf{N}_+,\ n \geq 2 \tag{1.5}$$

with $\tau^0\ x\ = x$ and $\tau^n\ x\ = \tau\ \tau^{n-1}\ x\ \ $.

For any $n \in \mathbf{N}_+$, writing: $x_1 \;\; = \dfrac{1}{x_1}$, $\;x_1, x_2, \ldots, x_n \;\; = \dfrac{1}{x_1 + \;\; x_2, \ldots, x_n}$, $n \geq 2$

for arbitrary indeterminates x_i, $1 \leq i \leq n$, we have:

$$x = \lim_{n \to \infty} \;\; a_1, \ldots, a_n \;\;, \; x \in \Omega \tag{1.6}$$

The probability structure of the sequence $\;\; a_n \;\; {}_{n \in \mathbf{N}_+}$ under λ, the Lebesgue measure, is described by the equations:

$$\lambda \;\; a_1 = i \;\; = \frac{1}{i \;\; i+1}, \; i \in \mathbf{N}_+ \tag{1.7}$$

$$\lambda \;\; a_{n+1} = i \,|\, a_1, \ldots, a_n \;\; := P_i \;\; s_n \;\;, \; i, n \in \mathbf{N}_+ \tag{1.8}$$

where:

$$s_n = \;\; a_n, \ldots, a_1 \tag{1.9}$$

and:

$$P_i \;\; x \;\; = \frac{x+1}{x+i \;\;\; x+i+1}, \; i \in \mathbf{N}_+, \; x \in I \tag{1.10}$$

Thus, under λ, the sequence $\;\; a_n \;\; {}_{n \in \mathbf{N}_+}$ is neither independent nor Markovian. Then, if B_I denotes the σ-algebra of Borel subsets of I, there is a probability measure γ on B_I:

$$\gamma \;\; A \;\; = \frac{1}{\log 2} \int_A \frac{dx}{x+1}, \; A \in B_I \tag{1.11}$$

called *Gauss' measure*, which makes $\;\; a_n \;\; {}_{n \in \mathbf{N}_+}$ into a strictly stationary sequence. Moreover, γ is τ-invariant, i.e.

$$\gamma \;\; \tau^{-1} \;\; A \;\;\; = \gamma \;\; A \;\;, \; A \in B_I \tag{1.12}$$

2 Random system with complete connections

Definition 2.1 A quadruple $\;\; W, W \;\;, \;\; X, X \;\;, u, P \;\;$ is named a homogeneous random system with complete connections (RSCC) if:

(i) $\;\; W, W \;\;$ and $\;\; X, X \;\;$ are arbitrary measurable spaces;

(ii) $u : W \times X \to W$ is a $\;\; W \;\; \otimes X \;\; , W \;\;$ - measurable function;

(iii) P is a transition probability function from $\;\; W, W \;\;$ to $\;\; X, X \;\;$.

Next, denote the element $\;\; x_1, x_2, \ldots, x_n \;\; \in X^n$ with $x^{\;n}$.

Definition 2.2 The functions $u^{\;n} \;\; : W \times X^n \to W$, $n \in \mathbf{N}_+$ are defined as follows:

$$u^{(n+1)}\left(\left(w,x^{(n+1)}\right)\right)=\begin{cases}u(w,x), & \text{if } n=0\\ u\left(u^{(n)}\left(\left(w,x^{(n)}\right)\right),x_{n+1}\right), & \text{if } n\in\mathbf{N}_+.\end{cases}\tag{2.1}$$

By convention, we will write $wx^{(n)}$ instead of $u^{(n)}\left(\left(w,x^{(n)}\right)\right)$.

Definition 2.3 The transition probability functions P_r, $r\in\mathbf{N}_+$ are defined by:

$$P_r(w,A)=\begin{cases}P(w,A), & \text{if } r=1,\\ \sum_{x_1\in X}P(w,x_1)\sum_{x_2\in X}P(wx_1,x_2)\dots\sum_{x_r\in X}P\left(wx^{(r-1)},x_r\right)\chi_A\left(x^{(r)}\right), & \text{if } r>1,\end{cases}\tag{2.2}$$

for any $w\in W$, $r\in\mathbf{N}_+$ and $A\in\mathcal{X}^r$, where χ_A is the characteristic function of the set A.

Definition 2.4 Assume that $X^0\times A=A$. Then we define:

$$P_r^n(w,A)=P_{n+r-1}\left(w,X^{n-1}\times A\right)\tag{2.3}$$

for any $w\in W$, $n,\ r\in\mathbf{N}_+$ and $A\in X^r$.

Theorem 2.5 (Existence theorem) Let $\{(W,\mathcal{W}),(X,\mathcal{X}),u,P\}$ be a homogeneous RSCC and let $w_0\in W$. Then there exist a probability space $\left(\Omega,\mathcal{K},P_{w_0}\right)$ and two chains of random variables $(\xi_n)_{n\in\mathbf{N}_+}$ and $(\zeta_n)_{n\in\mathbf{N}}$ defined on Ω with values in X and W respectively, such that:

(i) (a) $P_{w_0}\left((\xi_n,\dots,\xi_{n+r-1})\in A\right)=P_r^n(w_0,A)$,

(b) $P_{w_0}\left((\xi_{n+m},\dots,\xi_{n+m+r-1})\in A\,\middle|\,\xi^{(n)}\right)=P_r^m\left(w_0\xi^{(n)},A\right)$ P_{w_0}-a.e.

(c) $P_{w_0}\left((\xi_{n+m},\dots,\xi_{n+m+r-1})\in A\,\middle|\,\xi^{(n)},\zeta^{(n)}\right)=P_r^m(\zeta_n,A)$ P_{w_0}-a.e.

for any $n,m,r\in\mathbf{N}_+$ and $A\in X^r$, where $\xi^{(n)}$, $\zeta^{(n)}$ denote the random vectors $(\xi_1,\dots,\xi_n)$ and $(\zeta_1,\dots,\zeta_n)$ respectively.

(ii) $(\zeta_n)_{n\in\mathbf{N}}$ is a homogeneous Markov chain with initial distribution concentrated in w_0 and with the transition operator U defined by:

$$Uf(w)=\sum_{x\in X}P(w,x)f(wx)\tag{2.4}$$

for any f real $\mathcal{W}$-measurable and bounded function.

Remark 2.6 Letting $m=r=1$ in (i)(b) we obtain:

$$P_{w_0}\left(\xi_{n+1}\in A\,\middle|\,\xi^{(n)}\right)=P\left(w_0\xi^{(n)},A\right)\quad P_{w_0}\text{-a.e.}\tag{2.5}$$

that is the conditioned distribution of ξ_{n+1} by the past depends actually by this, through $u^{(n)}$. This fact justifies the name of *chain of infinite order* or *chain with complete connections* used for $(\xi_{n+1})_{n\in\mathbf{N}_+}$.

Remark 2.7 On account of (2.4) we have:

$$U^n f(w) = \sum_{x^{(n)}\in X^n} P_n(w, x^{(n)})\, f(wx^{(n)}),\ n\in\mathbf{N}_+ \tag{2.6}$$

for any f real W-measurable and bounded function.

Remark 2.8 The transition probability function of the Markov chain $(\zeta_n)_{n\in\mathbf{N}_+}$ is

$$Q(w,A) = \sum_{x\in X} P(w,x)\,\chi_A(wx) = P(w, A_w) \tag{2.7}$$

where $A_w = \{x\in X : wx\in A\}$, $w\in W$. It follows that the transition probability after n paths of the Markov chain $(\zeta_n)_{n\in\mathbf{N}}$ is:

$$Q^n(w,A) = P_n(w, A_w^n), \tag{2.8}$$

where $A_w^n = \{x^{(n)}\in X : wx^{(n)}\in A\}$

Definition 2.9 Let Q_n be the transition probability function defined by:

$$Q_n(w,A) = \frac{1}{n}\sum_{k=1}^{n} Q^k(w,A) \tag{2.9}$$

for any $w\in W$ and $A\in W$.

Definition 2.10 Let U_n be the Markov operator associated with Q_n. Then

(i) If there exists a linear bounded operator U^∞ from $L(W)$ to $L(W)$ such that:

$$\lim_{n\to\infty}\left\|U_n f - U^\infty f\right\| = 0, \tag{2.10}$$

for any $f\in L(W)$ with $\|f\|=1$, we say that U is *ordered*.

(ii) If:

$$\lim_{n\to\infty}\left\|U^n f - U^\infty f\right\| = 0, \tag{2.11}$$

for any $f\in L(W)$ with $\|f\|=1$, we say that U is *aperiodic*.

(iii) If U is ordered and $U^\infty(L(W))$ is one-dimensional space, then U is named *ergodic* with respect to $L(W)$.

(iv) If U is ergodic and aperiodic, then U is named *regular* with respect to $L(W)$ and the corresponding Markov chain has the same name.

Definition 2.12 If $((W,\mathcal{W}),(X,\mathcal{X}),u,P)$ is a RSCC which satisfies the properties

(i) (W,d) is a metric separable space;

(ii) $r_1 < \infty$, where:

$$r_k = \sup_{w' \neq w''} \sum_{X^k} P_k(w, x^{(k)}) \frac{d(w'x^{(k)}, w''x^{(k)})}{d(w', w'')}, \; k \in \mathbf{N}_+ \tag{2.12}$$

(iii) $r_1 < \infty$, where:

$$R_1 = \sup_{A \in \mathcal{X}} \sup_{w' \neq w''} \frac{\left| P(w', A) - P(w'', A) \right|}{d(w', w'')} \tag{2.13}$$

(iv) there exists $k \in \mathbf{N}_+$ such that $r_k < 1$

where $d(x,y) = |x-y|$, for any $x, y \in I$, then we say that this RSCC is a *RSCC with contraction*.

Theorem 2.13 Let (W,d) be a compact space and $((W,\mathcal{W}),(X,\mathcal{X}),u,P)$ a RSCC with contraction. The Markov chain associated to the RSCC is regular, if and only if, there exists a point $\tilde{w} \in W$ such that:

$$\lim_{n \to \infty} d(\sigma_n(\tilde{w}), w) = 0 \tag{2.14}$$

for any $w \in W$, where $\sigma_n(w) = \operatorname{supp} Q^n(w, \cdot)$, where $\operatorname{supp} \mu$ denotes the support of the measure μ.

Lemma 2.14 For any $m, n \in \mathbf{N}$, $w \in W$, we have:

$$\sigma_{m+n}(w) = \overline{\bigcup_{w' \in \sigma_m(w)} \sigma_n(w')} \tag{2.15}$$

where the line designates the topological aderence.

Definition 2.15 Let $((W,\mathcal{W}),(X,\mathcal{X}),u,P)$ be a RSCC. The RSCC is called *uniformly ergodic* if for any $r \in \mathbf{N}_+$ there exists a probability P_r^∞ on $\mathcal{X}^r$ such that $\lim_{n \to \infty} \varepsilon_n = 0$, where:

$$\varepsilon_n = \sup_{\substack{w \in W, r \in \mathbf{N}_+ \\ A \in \mathcal{X}^r}} \left| P_r^n(w, A) - P_r^\infty(A) \right|.$$

Theorem 2.16 Let (W,d) be a compact space. If the RSCC $((W,\mathcal{W}),(X,\mathcal{X}),u,P)$ with contraction has regular associated Markov chain, then it is uniform ergodic.

3 The Gauss–Kuzmin type theorem

Proposition 3.1 The function $P(x,i) = P_i(x)$ from (1.10) defines a transition probability function from (I, B_I) to $(\mathbf{N}, P(\mathbf{N}))$.

Proof. We have to verify that $\sum_{i\in\mathbf{N}} P(x,i) = 1$ for all $x \in I$. Since:

$$P(x,i) = (x+1)\left(\frac{1}{x+i} - \frac{1}{x+i+1}\right)$$

then:

$$\sum_{i\in\mathbf{N}} P(x,i) = (x+1)\cdot\frac{1}{x+1} = 1$$

Definition 3.2 Proposition 3.1 and relations (1.8) and (1.9) allows us to consider the random system with complete connections $\left((W,\mathcal{W}),(X,\mathcal{X}),u,P\right)$, with:

$$W = I,\ \mathcal{W} = B_I,\ X = \mathbf{N}_+,\ \mathcal{X} = P(\mathbf{N}_+)$$

and:

$$u(x,i) = \frac{1}{x+i},\ P_i(x) = \frac{x+1}{(x+i)(x+i+1)},\ x \in I,\ i \in \mathbf{N}_+$$

Remark 3.3 For $x_0 = 0$ and $P_0 = \lambda$ (see Theorem 2.5), the sequences $(\xi_n)_{n\in\mathbf{N}_+}$ and $(\zeta_n)_{n\in\mathbf{N}}$ associated with this RSCC coincide with the sequences $(a_n)_{n\in\mathbf{N}_+}$ and $(s_n)_{n\in\mathbf{N}}$, $s_0 = 0$ defined in (1.5) and (1.9).

Proposition 3.4 The RSCC from Definition 3.1 is a RSCC with contraction and its associated Markov operator U is regular with respect to $L(I)$ (the collection of all Lipschitz functions).

Proof. We have to verify the conditions from Definition 2.12. We have:

$$\frac{d}{dx}P(x,i) = \frac{i^2 - i - (x+1)^2}{(x+i)^2(x+i+1)^2}$$

and:

$$\frac{d}{dx}u(x,i) = -\frac{1}{(x+i)^2}$$

Hence, for any $x \in I$ and $i \in \mathbf{N}_+$, we have:

$$\sup_{x\in I}\left|\frac{d}{dx}P(x,i)\right| < \frac{1}{i^2}$$

and:

$$\sup_{x\in I}\left|\frac{d}{dx}u(x,i)\right| < \frac{1}{i^2}$$

Thus, $R_1 < \infty$ and $r_1 < \infty$. To proof the regularity of U with respect to $L(I)$, let us define recursively $x_{n+1} = \frac{1}{x_n + 2}$, $n \in \mathbf{N}_+$, with $x_0 = x$. Clearly, $x_{n+1} \in \sigma_1(x_n)$ and therefore Lemma 2.14 and an induction argument lead to the conclusion that $x_n \in \sigma_n(x_n)$, $n \in \mathbf{N}_+$. But $\lim_{n\to\infty} x_n = \sqrt{2} - 1$ for any $x \in I$. Hence:

$$d\left(\sigma_n(x), \sqrt{2}-1\right) \le \left|x_n - \left(\sqrt{2}-1\right)\right| \to 0,\ n \to \infty$$

Finally, the regularity of U with respect to $L(I)$ follows from Theorem 2.13.

Now, by the virtue of Theorem 2.16, the RSCC from Definition 3.1 is uniform ergodic. Moreover, $Q^n(\cdot,\cdot)$ converges uniformly to a probability measure Q^∞ and that there exists two positive constants $q < 1$ and k such that:

$$\left\|U^n f - U^\infty f\right\|_L \le kq^n \|f\|_L,\ n \in \mathbf{N}_+,\ f \in L(I) \tag{3.1}$$

where:

$$U^n f(\cdot) = \int_I f(y)\, Q^n(\cdot, dy),\ U^\infty f = \int_I f(y)\, Q^\infty(dy) \tag{3.2}$$

and $\|\cdot\|_L$ is the norm over $L(I)$:

$$\|f\|_L = \sup_{x\in I}|f(x)| + \sup_{x' \ne x''} \frac{|f(x') - f(x'')|}{|x' - x''|}$$

Proposition 3.5 The probability Q^∞ coincides with the Gauss' measure γ defined in (1.11).

Proof. By the virtue of uniqueness of Q^∞ we have to prove that:

$$\int_I Q(x, A)\, \gamma(dx) = \gamma(A),\ A \in B_I \tag{3.3}$$

Since the intervals $(0,u) \subset (0,1)$ generate B_I, it is suffices to check the above equation just for $A = (0,u)$, $0 < u \le 1$. We have:

$$Q(x, (0,u)) = \sum_{i: u(x,i) \in A} P(x,i),\ x \in I,\ A \in B_I$$

then:

$$Q(x, (0,u)) = \sum_{i \ge \left\lfloor \frac{1}{u} - x \right\rfloor + 1} P(x,i) = \frac{x+1}{x + \left\lfloor \frac{1}{u} - x \right\rfloor + 1}$$

thus:

$$\int_0^1 Q(x,(0,u))\gamma(dx) = \frac{1}{\log 2}\int_0^1 \frac{dx}{x+\left\lfloor \frac{1}{u}-x \right\rfloor+1} = \frac{1}{\log 2}\left(\int_0^{u^{-1}-\lfloor u^{-1}\rfloor} \frac{dx}{x+\left\lfloor \frac{1}{u} \right\rfloor+1} + \int_{u^{-1}-\lfloor u^{-1}\rfloor}^{1} \frac{dx}{x+\left\lfloor \frac{1}{u} \right\rfloor} \right)$$

$$= \frac{\log(1+u)}{\log 2} = \gamma((0,u)).$$

Proposition 3.6 Let μ be an arbitrary non–atomic probability measure on B_I. If

$$F_n(x) = F_n(x,\mu) = \mu(\tau^n < x),\ x \in I,\ n \in \mathbf{N} \tag{3.4}$$

with $F_0(x) = \mu((0,x))$, then for any $n \in \mathbf{N}$, F_n satisfies the following Gauss – Kuzmin type equation:

$$F_{n+1}(x) = \sum_{i\in\mathbf{N}_+} \left(F_n\left(\frac{1}{i}\right) - F_n\left(\frac{1}{x+i}\right) \right),\ x \in I. \tag{3.5}$$

Proof. Since $\tau^n = \dfrac{1}{a_{n+1}+\tau^{n+1}}$, it follows that:

$$F_n(x) = \mu(\tau^n < x) = \sum_{i\in\mathbf{N}_+} \mu\left(\frac{1}{a_{n+1}+x} < \tau^n < \frac{1}{a_{n+1}} \right) =$$

$$= \sum_{i\in\mathbf{N}_+} \left(F_n\left(\frac{1}{i}\right) - F_n\left(\frac{1}{x+i}\right) \right).$$

Assuming that for some $m \in \mathbf{N}$, the derivative F'_m exists everywhere in I and is bounded, it is easy to see by induction that F'_{m+n} exists and is bounded for all $n \in \mathbf{N}_+$. This allows us to differentiate (3.5) term by term, obtaining:

$$F'_{n+1} = \sum_{i\in\mathbf{N}} \frac{1}{(x+i)^2} F'_n\left(\frac{1}{x+i}\right) \tag{3.6}$$

Further, write $f_n(x) = (x+1)F'_n$, $x \in I$, $n \in \mathbf{N}$. Then (3.6) becomes $f_{n+1} = Uf_n$, $n \geq m$, with U being the linear operator defined as:

$$Uf(x) = \sum_{i\in\mathbf{N}_+} P_i(x) f\left(\frac{1}{x+i}\right) \tag{3.7}$$

Now, let μ be a probability measure on B_I such that $\mu \ll \lambda$. Then it can be shown that

$$\mu(\tau^{-n}(A)) = \int_A \frac{U^n f_0(x)}{x+1}dx,\ n \in \mathbf{N},\ A \in B_I \tag{3.8}$$

where:

$$f_0(x) = (x+1)F'_0(x),\ x \in I \tag{3.9}$$

with $F_0' = \frac{d\mu}{d\lambda}$.

Theorem 3.7 (Gauss – Kuzmin Theorem) Let μ be a probability measure on B_I such that $\mu \ll \lambda$. If the density F_0' of μ is a Riemann integrable function, then:

$$\lim_{n\to\infty} \mu\left(\tau^n < x\right) = \frac{1}{\log 2}\log\left(x+1\right),\ x \in I \tag{3.10}$$

If the density F_0' of μ is a Lipschitz function, then there exist two positive constants $q<1$ and k such that for all $x \in I$ and $n \in \mathbf{N}_+$:

$$\mu\left(\tau^n < x\right) = \frac{1}{\log 2}\left(1+\theta q^n\right)\log\left(x+1\right) \tag{3.11}$$

where $\theta = \theta\left(\mu, n, x\right)$, with $|\theta| \le k$

Proof. Let F_0' be a Lipschitz function. Then, $f_0 \in L\left(I\right)$ and by the virtue of (3.2):

$$U^\infty f_0 = \int_I f_0\left(x\right) Q^\infty\left(dx\right) = \int_0^1 f_0\left(x\right)\gamma\left(dx\right) = \int_0^1 F_0'\left(x\right)dx = \frac{1}{\log 2} \tag{3.12}$$

According to (3.1) there exist two constants $q<1$ and k such that:

$$U^n f_0 = U^\infty f_0 + T^n f_0,\ n \in \mathbf{N}_+ \tag{3.13}$$

with $\left\|T^n f_0\right\|_L \le kq^n$. Further, consider $C\left(I\right)$ the metric space of real continuous functions defined on $\left[0,1\right]$ with the supremum norm. Since $L\left(I\right)$ is a dense subset of $C\left(I\right)$ we have:

$$\lim_{n\to\infty}\left|T^n f_0\right| = 0 \tag{3.14}$$

for $f_0 \in C\left(I\right)$. Therefore (3.14) is valid for measurable f_0 which is γ-almost surely continuous, that is for Riemann integrable f_0. Thus:

$$\lim_{n\to\infty}\mu\left(\tau^n < x\right) = \lim_{n\to\infty}\int_0^x U^n f_0\left(u\right)\frac{1}{u+1}du = \int_0^x U^\infty f_0\left(u\right)\frac{1}{u+1}du =$$

$$= \frac{1}{\log 2}\log\left(u+1\right)\Big|_0^x = \frac{1}{\log 2}\log\left(x+1\right).$$

Equation (3.11) is equivalent with:

$$F_n\left(x\right) = \left(1+\theta q^n\right)F_\infty\left(x\right),\ n \in \mathbf{N}_+$$

which results from:

$$U^n f_0\left(x\right) = \left(1+\theta q^n\right)U^\infty f_0\left(x\right),\ n \in \mathbf{N}_+,\ x \in I$$

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SOUND ABSORBING MATERIALS

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Abstract: *Nowadays, sound absorbing materials are used on a large scale and in different industries as well as for people's comfort in homes and other public areas. Lot of research is about improving the absorptive and reflective characteristics of the materials.*

Keywords*: sound, absorbing materials, ship*

In the present, there is a variety of sound absorbing materials that are widely used for the control of noise in a variety of different situations. These materials are used to control reverberant sound generated by machinery and other equipment (to reduce reflections from surfaces and to decrease reverberation within spaces), to increase the sound attenuation in air-conditioning ducts, to reduce the sound radiated from the machine and other applications. The function of absorptive materials is to transform the impinging sound energy into heat.

Sound absorbing materials exist in many different forms. The most common types are Rockwool, fiberglass, polyurethane and cellulose fibers. Also, there are other materials which include:

- curtains and drapes
- carpets
- thin porous sheets, often mounted on a honeycomb structure
- hanging baffles used to reduce reverberation in factories, and other enclosed spaces
- felt materials
- hollow concrete blocks with a small opening to the outside - to create a Helmholtz resonator

Sound absorbing materials are usually fibrous, lightweight and porous.

Porous materials are characterized by having an open structure that is accessible by the air. Thus, air can be pressed through the material more or less easily depending on the flow resistance. The absorption properties are caused by viscous friction between the moving air molecules in the sound waves and the often huge internal surface area of the structure whereby the (kinetic) sound energy is converted into heat.

Porous materials are frequently fragile, and, as a result, it is necessary to protect the exposed surface of the material.

Typical protective surfaces include:

- thin impervious membranes of plastic or other material,
- perforated facings of metal, plastic, or other material,
- fine wire-mesh screens,
- sprayed-on materials such as neoprene,
- thin porous surfaces.

For industrial applications, wire meshes and perforated metal or hardboard is most practical. A perforated covering will have little effect on the absorption efficiency of the materials if it has only trivial airflow resistance. The perforated area should be at least 10 or 20% of the total area covered and the distance between openings must not be more than one-quarter wavelength at the highest frequency of interest (about 1 cm, for example, at 8,000 Hz). Impermeable coverings, which may be required to prevent absorption of oil, water or solvent, must have a very low surface density (less than 0.1 kg/m2). Mylar, polyester or plastic films, 0.025 mm thick, or less, are commonly used, sometimes completely enclosing the absorbent material. Although this may slightly reduce the high-frequency absorption, it will improve the low-frequency absorption. In general, the protective facings or porous materials should not be painted. If it is painted, it must be done with a very thin coat that will not reduce the ability of the facing to transmit a sound wave.

Mineral wool consists of thin fibers pressed and glued together. The fibers are made from melted glass (Glasswool) or stone (Rockwool) much like “Candy Floss”. Mineral wool is used as porous sound absorbers, very often in the form of tiles that can be mounted in a suspended ceiling system. Such ceilings will often be placed below ventilation ducts and other technical installations whereby a large distance (typically between 20 cm and one meter) is ensured to the hard concrete deck behind. Hereby the ceiling can absorb efficiently over a wide frequency range – as well as hide the installations. Mineral wool ceiling tiles are normally given a carefully controlled layer of special paint from the factory to make them look like normal (white) plaster ceilings as much as possible. However, if one tried to repaint them, the porous properties and so the absorption normally disappears.

The next examples apply to ceilings and walls. Some of these solutions for the rooms and other spaces inside a building can be adopted aboard a ship. Figure 1 shows the example of a wood slat panel treatment that effectively screens the acoustic blanket. This solution isn’t recommended onboard a ship because wood if a fire hazard. Another example are perforated metal panels, as shown in figure 2. For best results, the material should be as thin as possible, with the smallest hole diameter and the greatest open area (the greatest number of holes). Some absorptive materials are attractively designed to be exposed to view, such as normal suspended ceiling tiles. Generally, thicker porous materials provide better sound absorption. 5/8-inch thick ceiling tiles have an NRC of 0.50 when mounted in a lay-in grid ceiling. A 1-inch thick glass fibre ceiling tile can have an NRC rating of 0.80 or greater. Figure 3 illustrates the appearance of a suspended acoustical tile ceiling. Another approach to adding acoustic absorption to the space is to suspend acoustic baffles as shown in figure 4.

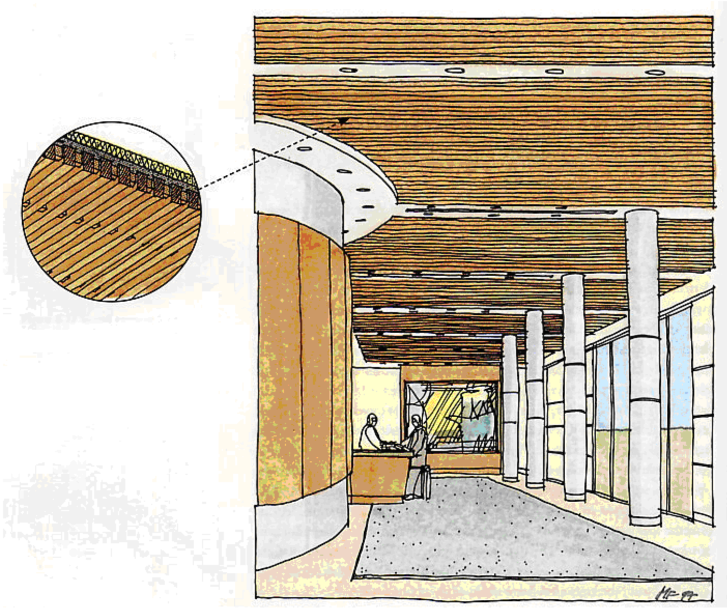


Fig. 1

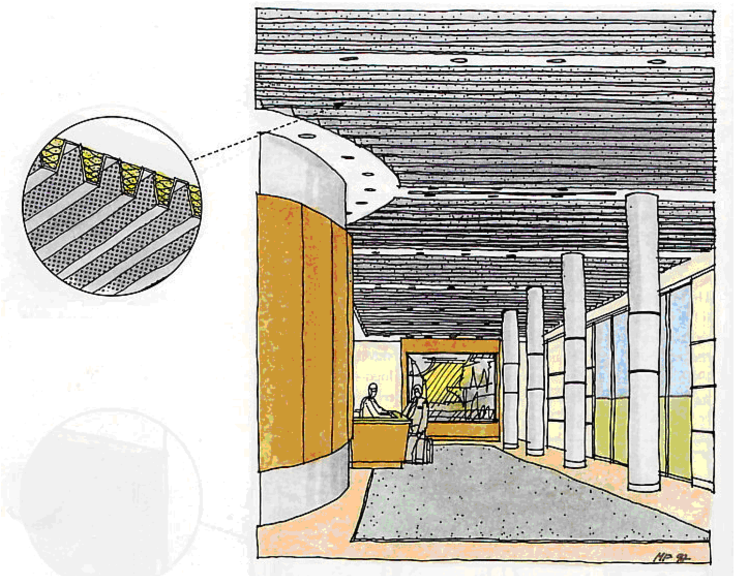


Fig. 2

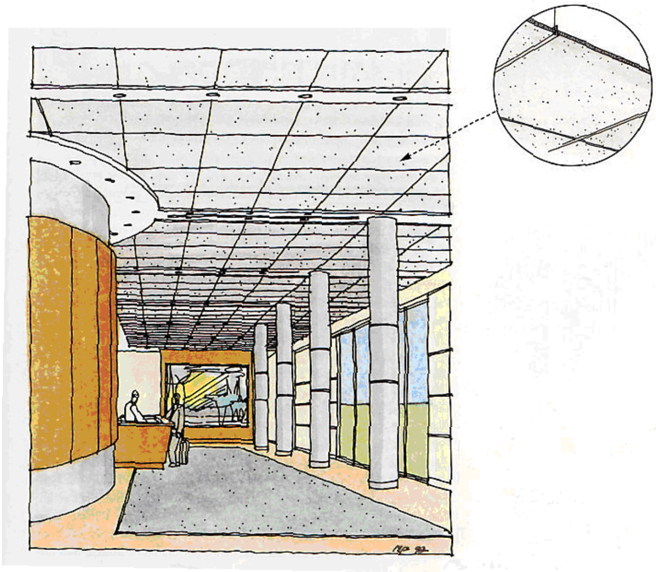


Fig. 3

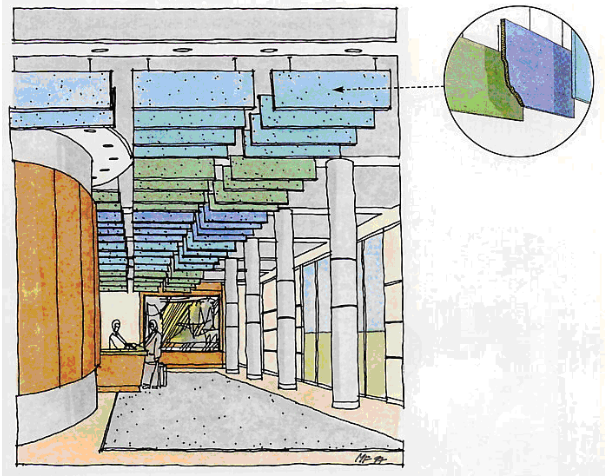


Fig. 4

Open-cell foam panels are effective sound absorbers because they have increased surface area due to the contoured surface of the foam. Figure 5 illustrates such an application, but again, it is not recommended onboard a ship.

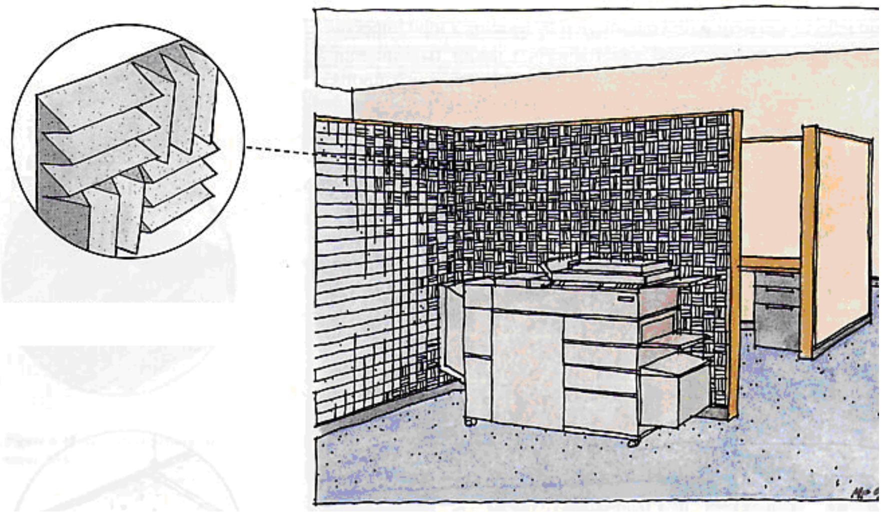


Fig. 5

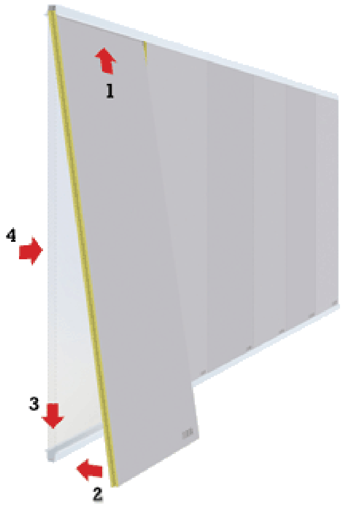
Aboard a ship, when it is not possible to decrease the noise characteristics of a system or the cost of reducing noise at its source is too high in terms of system performance or overall expense, isolation or insulation may provide a viable noise mitigation alternative. The source can be isolated from the structure or the crew can be isolated from the transmitted noise. The classic treatment to noise control is the addition of insulation to absorb/attenuate acoustic energy. Also, one need to consider other characteristics of the material used onboard a ship: fire proof of the material, stress resistance (variations of humidity, temperature, vibrations), the weight, adhesive power.

•**Acoustic Absorption.** Absorptive treatments reduce noise in all accommodation compartments or spaces that house loud machinery. They absorb noise that would have been reflected through applied surfaces. This treatment is very important for ships, because most surfaces can be reflective. Materials such as absorbent ceiling panels, floor carpeting, drapes, or special absorbent wall coverings will reduce noise by reducing the reflected sound. Protective shielding for the absorbing material should generally be perforated. Acoustic absorptive treatments are only effective for attenuating the noise within a space by preventing reflections; they do little for reducing airborne or structure borne noise generated in and transmitted from other compartments.

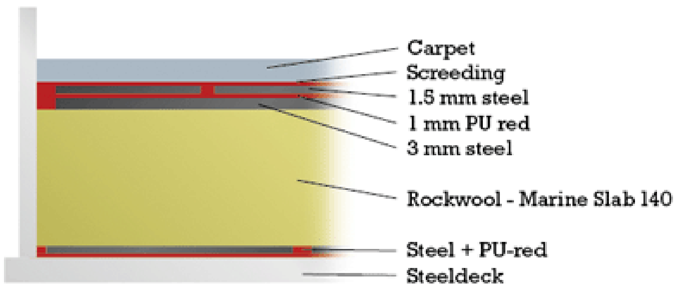
Absorption: 2 x 50 Rockwool Marine Slab 80

2 x 50 mm Rockwool Marine Slab 80
during testing of absorption value

• **Acoustic Insulation**. This treatment is applied to provide sound insulation for an area or a compartment. An example is the application of acoustic sound insulation for control booths in the propulsion spaces. Insulation treatments take various forms such as rigid or semi rigid boards and blankets depending on the application. Some versions can be layered with a noise barrier material, technically referred to as a "limp mass" sandwiched between fiberglass or foam. The limp mass helps block noise transmission though the material.



• **Floating Floor/Floating Room Treatments.** Instead of isolating the machinery, these treatments isolate the compartment. These treatments vary in details of composition, but fundamentally consist of a dense rigid false floor, isolated from the structural deck by fibrous glass batts or resilient supports. For a floating room, false bulkheads are attached to the false floor and are only connected to ship structure by flexible supports at the top. Acoustic absorptive material, such as fibrous glass, is attached to the structural bulkheads behind the false bulkheads. Floating floor treatments are often used for engineers' control rooms or audiometric test booths aboard ships.



• **Advanced Acoustic Materials**

- **Viscoelastic Laminated Sheet Metal**. Viscoelastic laminated sheet metal has a sandwich layer of damping material that can be used to control noise transmission by damping the structure and removing energy from the resonant vibration. The Viscoelastic laminated sheet can be applied to exterior surfaces or placed between surfaces of structures or equipment.


- Viscoelastic Polymer Coating. This soundproof coating material has been incorporated into *Sea Fighter*, the US Office of Naval Research’s experimental craft for the US Navy’s Littoral Combat Ship. The soundproof coating material is applied to the hull and mission bay of *Sea Fighter* to reduce vibration and noise. According to its manufacturer, the soundproof coating material helps to reduce noise on the *Sea Fighter* by 15 decibels.

• **Acoustic Pipe Insulation**

Generally, any material applied over the surface of a pipe will provide some sound attenuation. The amount of attenuation depends on the type of materials used, the thickness and the method of application.

Acoustic insulation does not reduce the sound energy produced within the valve or pipe. Rather, it absorbs or blocks the transmission of a portion of the sound energy, reducing the external noise level. The noise created by a valve or other restrictions in the pipeline can be transmitted for relatively long distances through the pipe wall and fluid stream. For this reason, it is often necessary to treat the entire piping system to eliminate the noise problem completely. Where the piping system passes through separate enclosures or rooms, it is often possible to reduce the noise level within a given room by insulating only that portion of the piping contained within the room.

Noise in compressible fluid systems generally predominates in the frequency range of 1000 to 8000 Hz. In this frequency range, the most effective acoustic materials are porous or fibrous materials such as fiberglass, Rockwool or foam materials. Fiberglass blankets are commonly used to wrap pipelines and are available for temperatures up to 500 ^{0}C. Generally, the attenuation of a given material increases with thickness, density and the frequency of the sound. For example, fiberglass has an attenuation of 6-8 dB/cm thickness at 1000 Hz and 11-13 dB/cm thickness at 8000 Hz; Rockwool has an attenuation of 4 dB/cm thickness at 1000 Hz and 9-13 dB/cm thickness at 8000 Hz. Values for other materials or densities are available from the manufacturers of the material. The values given above are applicable over a limited range of thickness (generally 5–10 cm.). Greater thickness generally will provide greater attenuation per inch at low frequencies and may provide lower attenuation per inch at high frequencies.

The attenuation values above are for insulation with no outer cover, or with a porous outer cover such as untreated cloth. A significant increase in attenuation can be achieved by using an impervious cover such as thin sheet steel, aluminum, lead, rubber or cloth coated with an airtight material. For maximum attenuation, at low frequencies (below 1000 Hz) the cover material should be dense and limp. Lead and lead-coated cloths are ideal in this respect. At higher frequencies the limpness and density are not as critical.

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ALGEBRAIC DOMAIN DECOMPOSITION METHODS

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Abstract: *By using the domain decomposition methodology, we construct several algebraic domain decomposition methods for certain algebraic systems with sparse matrix. These methods are highly parallelizable. We show that these methods are convergent and we also discuss the eigenvalue distributions of the corresponding iterative matrices in order to analyze the convergence factors of these methods.*

Keywords: *algebraic systems, decomposition*

1.1 Algebraic domain decomposition methods

Let's consider the linear algebraic system:

$$Au = f \tag{1.1}$$

where matrix A is a o $(p-1)x(p-1)$ block square matrix, denotes as:

$$A = \begin{bmatrix} A_{1,1} & \cdots & A_{1,2p-1} \\ \vdots & & \\ A_{2p-1,1} & \cdots & A_{2p-1,2p-1} \end{bmatrix}, \quad u = \begin{bmatrix} u_1 \\ \vdots \\ u_{2p-1} \end{bmatrix}, \quad f = \begin{bmatrix} f_1 \\ \vdots \\ f_{2p-1} \end{bmatrix}$$

As in the domain decomposition methods, we partition the unknown vector u into p new subvectors with overlapping. We let $\tilde{u}$ denoted as a vector defined by these p new sub-vectors:

$$\tilde{u} = \begin{bmatrix} x_1 \\ \vdots \\ x_p \end{bmatrix}, \text{ with } x_1 = \begin{bmatrix} u_1 \\ u_2 \end{bmatrix}, \quad x_p = \begin{bmatrix} \tilde{u}_{2p-2} \\ u_{2p-1} \end{bmatrix}, \quad x_i = \begin{bmatrix} \tilde{u}_{2i-2} \\ u_{2i-1} \\ u_{2i} \end{bmatrix} \tag{1.2}$$

with $2 \le i \le p-1$. Note that $\tilde{u}_{2i}$ are the unknown vectors associated with the overlapping to the subvector $\tilde{u}_{2i}=u_{2I}$ $i=1,...,p$. In the same way, we can introduce a new vector $\tilde{f}$ from righthandside vector f. Thus, the matrix A is divided into $p\times p$ corresponding block submatrices with overlapping. Then, a corresponding new matrix $\tilde{A}$ can be defined by these $p\times p$ block submatrices:

$$\tilde{A} = \begin{bmatrix} \tilde{A}_{1,1} & \cdots & \tilde{A}_{1,p} \\ \vdots & & \\ \tilde{A}_{p,1} & & \tilde{A}_{p,p} \end{bmatrix} \tag{1.3}$$

where:

$$\tilde{A}_{1,1} = \begin{bmatrix} A_{1,1} & A_{1,2} \\ A_{2,1} & A_{2,2} \end{bmatrix}, \quad \tilde{A}_{p,p} = \begin{bmatrix} A_{2p-2,2p-2} & A_{2p-2,2p-1} \\ A_{2p-1,2p-2} & A_{2p-1,2p-1} \end{bmatrix}$$

$$\tilde{A}_{1,p} = \begin{bmatrix} 0 & A_{1,2p-1} \\ 0 & A_{2,2p-1} \end{bmatrix}, \quad \tilde{A}_{1,i} = \begin{bmatrix} 0 & A_{1,2i-1} & A_{1,2i} \\ 0 & A_{2,2i-1} & A_{2,2i} \end{bmatrix} \quad 2 \le i \le p-1$$

$$\tilde{A}_{p,1} = \begin{bmatrix} A_{2p-2,1} & 0 \\ A_{2p-1,1} & 0 \end{bmatrix}, \quad \tilde{A}_{p,i} = \begin{bmatrix} A_{2p-2,2i-2} & A_{2p-2,2i-1} & 0 \\ A_{2p-1,2i-2} & A_{2p-1,2i-1} & 0 \end{bmatrix} \quad 2 \le i \le p-1$$

$$\tilde{A}_{i,1} = \begin{bmatrix} A_{2i-2,1} & 0 \\ A_{2i-1,1} & 0 \\ A_{2i,1} & 0 \end{bmatrix} \quad \tilde{A}_{i,p} = \begin{bmatrix} 0 & A_{2i-2,2p-1} \\ 0 & A_{2i-1,2p-1} \\ 0 & A_{2i,2p-1} \end{bmatrix} \quad 2 \le i \le p-1$$

$$\tilde{A}_{i,j} = \begin{bmatrix} 0 & A_{2i-2,2j-1} & A_{2i-2,2j} \\ 0 & A_{2i-1,2j-1} & A_{2i-1,2j} \\ 0 & A_{2i,2j-1} & A_{2i,2j} \end{bmatrix} \quad 2 \le i < j \le p-1$$

$$\tilde{A}_{j,i} = \begin{bmatrix} A_{2i-2,2j-2} & A_{2i-2,2j-1} & 0 \\ A_{2i-1,2j-2} & A_{2i-1,2j-1} & 0 \\ A_{2i,2j-2} & A_{2i,2j-1} & 0 \end{bmatrix} \quad 2 \le i < j \le p-1$$

$$\tilde{A}_{i,i} = \begin{bmatrix} A_{2i-2,2i-2} & A_{2i-2,2i-1} & A_{2i-2,2i} \\ A_{2i-1,2i-2} & A_{2i-1,2i-1} & A_{2i-1,2i} \\ A_{2i,2i-2} & A_{2i,2i-1} & A_{2i,2i} \end{bmatrix} \quad 2 \le i \le p-1$$

From this definition, we obtain a new linear system:

$$\hat{A}\hat{u} = \hat{f} \tag{1.4}$$

which is associated with system (1.1). The following theorem gives the relation between equation (1.1) and system (1.4).

Theorem 1.1.1 Suppose that $A_{2i,2i}$ are nonsingular for all $i = 1, \ldots, p-1$. Then, the solution of system (1.1) can be constructed from the solution of the problem (1.4) and vice versa.

Proof Let *u* the solution of system (1.1). We can construct a new vector $\tilde{u}$ as above from u by letting the overlapping. Then, this $\tilde{u}$ is the solution of system (1.4).

Suppose that $\tilde{u}$ is the solution of problem (1.4). Because $A_{2i,2i}$ are nonsingular for $i = 1, \ldots, p-1,$ we have $u_{2i} = \tilde{u}_{2i}, \quad i = 1, \ldots, p-1$ from the equation $A_{2i,2i}\left(u_{2i} - \tilde{u}_{2i}\right) = 0$. A direct consequence is that the vector $u = \left(u_1^T, \ldots, u_{2p-1}^T\right)$ is the solution of system (1.1).

Theorem 1.1.1 implies that if the solution of (1.1) is unique, the problem (1.4) has only one solution.

Let us denote $\lambda(A)$ to be the set of eigenvalues of matrix *A*.

Theorem 1.1.2

$$\lambda(\tilde{A}) = \lambda(A) \cup \left(\bigcup_{i=1}^{p-1} \lambda\{A_{2i,2i}\}\right)$$

Proof We first prove $\lambda(\tilde{A}) \subseteq \lambda(A) \cup \left(\bigcup_{i=1}^{p-1} \lambda\{A_{2i,2i}\}\right)$. From $\tilde{A}\tilde{u} = \tilde{\lambda}\tilde{u}$ we have $A_{2i,2i}(u_{2i} - \tilde{u}_{2i}) = \tilde{\lambda}(u_{2i} - \tilde{u}_{2i})$ $i = 1,\ldots,p-1$. If $u_{2i} \neq \tilde{u}_{2i}$ for some $i \in \{1,\ldots,p-1\}$, then $\tilde{\lambda} \in \lambda(A_{2i,2i})$. Otherwise, if $u_{2i} = \tilde{u}_{2i}$ $(\forall i \in \{1,\ldots,p-1\})$, then $\tilde{\lambda} \in \lambda\{A\}$.

Now we prove that $\lambda(A) \cup \left(\bigcup_{i=1}^{p-1} \lambda\{A_{2i,2i}\}\right) \subset \lambda(\tilde{A})$. From $Au = \lambda u$ we have $\tilde{A}\tilde{u} = \lambda\tilde{u}$ with $\tilde{u}_{2i} = u_{2i}$ as in (1.2). If $A_{2i,2i}v_{2i} = \lambda v_{2i}$ for some $i \in \{1,\ldots,p\}$, then we can construct a vector $\tilde{u}$ in the following way such that $\tilde{A}\tilde{u} = \lambda\tilde{u}$. Let $w = -(0,\ldots 0, v_{2i}^T A_{2i+1,2i}^T, \ldots, v_{2i}^T A_{2p-1,2i})$. If this $\lambda \notin \lambda(A)$ then the equation $(A - \lambda I)u = w$ has only one solution *u*. By setting:

$$\tilde{u}_{2j} = \begin{cases} u_{2j} & j \neq i \\ u_{2i} + v_{2j} & j = i \end{cases}$$

For this $\tilde{u}$, we can easily show that $\tilde{A}\tilde{u} = \lambda\tilde{u}$. Hence, $\lambda(A) \cup \left(\bigcup_{i=1}^{p-1} \lambda\{A_{2i,2i}\}\right) \subset \lambda(\tilde{A})$. □

Before giving a definition an asynchronous Schwarz algorithm for problem (1.1), we first, decompose linear system (1.1) into *p* subproblems: to find $x*_i$ such that:

$$\tilde{A}_{i,i} x*_i = \tilde{f}_i - \sum_{j \neq i} \tilde{A}_{i,j} x_j, \quad 1 \leq i \leq p \tag{1.5}$$

where $\tilde{u} = (x_1^T, \ldots, x_p^T)^T$ is regarded as a known vector and $\tilde{u}* = (x*_1^T, \ldots, x*_p^T)^T$ as an unknown vector. Let φ_i denote as a solver (1.5). This φ_i is either direct solver or iterative solver. Now we list several possible choices φ_i as the examples:

a) $\tilde{u}* = \varphi_i(\tilde{u})$ and $\tilde{A}_{i,i}\tilde{u}* = \tilde{f}_i - \sum_{j \neq i} \tilde{A}_{i,j} x_j$

b) $\tilde{u}* = \varphi_i(\tilde{u})$ is defined by:

$$r_i = \tilde{f}_i - \sum_{1 \leq j \leq p} \tilde{A}_{i,j} x_j$$

$$\alpha_i = (r_i, r_i)/(A_{i,i} r_i, r_i) \quad x*_i = x_i + \alpha_i r_i$$

c) let $\widetilde{A}_{i,i} = M_i - N_i$ where M_i is an invertible matrix. Then,

$$\widetilde{u}^* = \varphi_i(\widetilde{u}) \text{ and } x^*_i = x_i + M_i^{-1}\left(\widetilde{f}_i - \sum_{1\le j\le p} \widetilde{A}_{i,j}x_j\right)$$

Let's distribute the computation on the machine with p processors. Each processor is assigned to solve on subproblem (1.5). Denote N as the whole positive integer set. If we let all processor keep on calculating by using the most recent available data from neighbour processors, then we have following asynchronous method:

Let $\widetilde{u}^{(0)}$ be a given initial guess vector. The vector sequence $\widetilde{u}^{(k)}$ will be defined by the recursion:

$$\begin{cases} x_i^{k+1} = \phi_i\left(x_1^{s_1(k)},\ldots,x_p^{s_p(k)}\right) \\ x_i^{(k+1)} = x_i^{s_i(k)} + \omega_i\left(x_i^{(k+1)} - x_i^{s_i(k)}\right) \text{ if } \; i \in J(k) \\ x_i^{(k+1)} = x_i^{(k)} \; \text{dacã } \; i \notin J(k) \\ \widetilde{u}^{(k+1)} = \left(x_1^{(k+1)},\ldots,x_p^{(k+1)}\right) \end{cases} \tag{1.6}$$

where $\{J(k)\}_{k\in N}$ is a sequence of nonempty subsets of $\{1,\ldots,p\}$. In fact $J(k)$ is the set of subvectors to be updated at step *k*. Here, $S = \{s_1(k),\ldots,s_p(k)\}$ is a sequence of elements of N^p with the following properties:

$$\begin{array}{l} s_i(k) \le k \quad (\forall)k \in N, \quad (\forall)i \in \{1,\ldots,p\} \\ \lim\limits_{k\to\infty} s_i(k) = \infty \quad (\forall)i \in \{1,\ldots,p\} \end{array}$$

Such procedure is called **chaotic relaxation Schwarz (CRS) algorithm** and identified by $(\phi_i, \widetilde{u}^{(0)}, J, S)$. Selecting the set $J(k)$ and set $S = \{s_1(k),\ldots,s_p(k)\}$ we give several special cases of CRS algorithm:

a) Algebraic Multiplicative Schwarz (AMS) Algorithm with:

$$\begin{array}{l} s_i(k) = k \\ J(k) = 1 + k \bmod(p), (\forall)k \end{array}$$

b) Algebraic Additive Schwarz additive (AAS) Algorithm with:

$$\begin{array}{l} s_i(k) = k \\ J(k) = \{1,\ldots,\mathrm{p}\} \quad (\forall)k \end{array}$$

1.2 Direct Subsolver for All Sub-problems

We use direct solver for the subproblems in our CRS method. In order to analyse the convergence factor, we discuss the eigenvalue distribution of iterative matrix of AMS and AAS methods. Let the matrix

$$\widetilde{A} = \widetilde{D} + \widetilde{L} + \widetilde{U}$$

where $\tilde{D}$ is a block diagonal matrix, $\tilde{L}$ is a block lower triangular matrix, and $\tilde{U}$ is a block upper triangular matrix:

$$\tilde{D} = \begin{bmatrix} \tilde{A}_{1,1} & & 0 \\ & \ddots & \\ 0 & & \tilde{A}_{p,p} \end{bmatrix}, \quad \tilde{L} = \begin{bmatrix} 0 & & & \\ \tilde{A}_{2,1} & 0 & & \\ \vdots & \ddots & \ddots & \\ \tilde{A}_{p,1} & \cdots & \tilde{A}_{p,p-1} & 0 \end{bmatrix} \quad \tilde{U} = \begin{bmatrix} 0 & \tilde{A}_{1,2} & \cdots & \tilde{A}_{1,p} \\ & \ddots & \ddots & \vdots \\ & & 0 & \tilde{A}_{p-1,p,} \\ & & & 0 \end{bmatrix}$$

Assume for $k = 1,2,\ldots$ we define a new sequence $y^{(k)}$ by:

$$y^{(k)} = \begin{bmatrix} x_1^{(k-1)p+1} \\ \vdots \\ x_p^{kp} \end{bmatrix} \quad \text{with } y^{(0)} = \tilde{u}^{(0)}$$

If p subproblems are all solved by the direct solver ϕ_i with $\omega_i = \omega$, the AMS method can be described in one simple form

$$\tilde{H}y^{(k+1)} = \tilde{B}y^{(k)} + \omega\tilde{f}$$

where $\tilde{H} = \tilde{D} + \omega\tilde{L}$ and $\tilde{B} = (1-\omega)\tilde{D} - \omega\tilde{U}$, and the AAS method can be rewritten as

$$\tilde{D}\tilde{u}^{(k+1)} = (\tilde{D} - \omega\tilde{A})\tilde{u}^{(k)} + \omega\tilde{f}$$

Then the iterative matrices of the AMS method and AAS method satisfy the following theorem.

Theorem 1.2.1 If $A_{i,2j} = 0$ for $|i-2j| \geq 2$, then

i) The iterative matrix of the AMS method satisfies:

$$\lambda\left(\tilde{H}^{-1}\tilde{B}\right) \subseteq \lambda\left((\dot{D}+\omega\dot{L})^{-1}((1-\omega)\dot{D}-\omega\dot{L})\right)$$

and:

$$\lambda\left((\dot{D}+\omega\dot{L})^{-1}((1-\omega)\dot{D}-\omega\dot{L})\right) \cup \{1-\omega\} \subseteq \lambda\left(\tilde{H}^{-1}\tilde{B}\right)$$

ii) The iterative matrix of the AAS method satisfies:

$$\lambda\left(\tilde{D}^{-1}(\tilde{D}-\omega\tilde{A})\right) \subseteq \lambda\left(\dot{D}^{-1}(\dot{D}-\omega\dot{A})\right)$$

and:

$$\lambda\left(\dot{D}^{-1}(\dot{D}-\omega\dot{A})\right) \cup \{1-\omega\} \subseteq \lambda\left(\tilde{D}^{-1}(\tilde{D}-\omega\tilde{A})\right)$$

Here the block diagonal matrix $\dot{D}$, lower triangular matrix $\dot{L}$, and upper triangular matrix $\dot{U}$ in the sum expression

$$\dot{A} = \dot{D} + \dot{L} + \dot{U}$$

have the forms:

$$\dot{D} = \begin{bmatrix} \dot{A}_{1,1} & & & \\ & \dot{A}_{3,3} & 0 & \\ & 0 & \ddots & \\ & & & \dot{A}_{2p-1,2p-1} \end{bmatrix} \qquad \dot{L} = \begin{bmatrix} 0 & & & & \\ \dot{A}_{3,1} & 0 & & & \\ \dot{A}_{5,1} & \dot{A}_{5,3} & & & \\ \vdots & \ddots & \ddots & & \\ \dot{A}_{2p-1,1} & & \dot{A}_{2p-1,2p-3} & 0 & \end{bmatrix}$$

$$\dot{U} = \begin{bmatrix} 0 & & \dot{A}_{1,3} \cdots & \dot{A}_{1,2p-1} \\ & \ddots & \cdots & \vdots \\ & & A_{2p-5,2p-3} & A_{2p-5,2p-1} \\ & & 0 & A_{2p-3,2p-1} \\ & & & 0 \end{bmatrix}$$

where:

$$\dot{A}_{1,2j-1} = A_{1,2j-1} - A_{1,2}A_{2,2}^{-1}A_{2,2j-1} \qquad \text{for } 1 \le j \le p$$

$$\dot{A}_{2p-1,2j-1} = A_{2p-1,2j-1} - A_{2p-1,2p-2}A_{2p-2,2p-2}^{-1}A_{2p-2,2j-1} \qquad \text{for } 1 \le j \le p$$

$\dot{A}_{2i-1,2j-1} = A_{2i-1,2j-1} - A_{2i-1,2i-2}A_{2i-2,2i-2}^{-1}A_{2i-2,2j-1} - A_{2i-1,2i}A_{2i,2i}^{-1}A_{2i,2j-1}$ for $1<i<p$ and $1 \le j \le p$

Proof Suppose that $\lambda \in \lambda(\tilde{H}^{-1}\tilde{B})$ and $\tilde{u}$ is the corresponding eigenvector, i.e. $\tilde{H}^{-1}\tilde{B}\tilde{u} = \lambda\tilde{u}$. So $\tilde{B}\tilde{u} = \lambda\tilde{H}\tilde{u}$. Let $\dot{u} = (u_1^T, \ldots, u_{2p-1}^T)^T$ be the vector defined by the subvectors of $\tilde{u}$. Then we have

$$\lambda(\dot{D} + \omega\dot{L})\dot{u} = ((1-\omega)\dot{D} - \omega\dot{U})\dot{u}.$$

Hence, $\lambda \in \lambda((\dot{D} + \omega\dot{L})^{-1}((1-\omega)\dot{D} - \omega\dot{U}))$. Thus,

$$\lambda(\tilde{H}^{-1}\tilde{B}) \subseteq \lambda((\dot{D} + \omega\dot{L})^{-1}((1-\omega)\dot{D} - \omega\dot{U}))$$

Let $\lambda((\dot{D} + \omega\dot{L})^{-1}((1-\omega)\dot{D} - \omega\dot{U}))$ and $\dot{u}$ be the corresponding eigenvector, i.e. $\lambda(\dot{D} + \omega\dot{L})^{-1}((1-\omega)\dot{D} - \omega\dot{U})\dot{u}$. Now we construct an eigenvector $\tilde{u}$ of $\tilde{H}^{-1}\tilde{B}$ from this eigenvector $\dot{u}$. Let the subvectors of $\tilde{u}$ be defined by the corresponding subvectors of $\dot{u}$. The other subvectors of $\tilde{u}$ are uniquely determined by the equation $\lambda\tilde{H}\tilde{u} = \tilde{B}\tilde{u}$ if $\lambda \ne 1-\omega$. It follows that $\lambda\{\tilde{H}^{-1}\tilde{B}\}$. Thus,

$$\lambda((\dot{D} + \omega\dot{L})^{-1}((1-\omega)\dot{D} - \omega\dot{U})) \setminus \{1-\omega\} \subseteq \lambda(\tilde{H}^{-1}\tilde{B}). \square$$

We rewrite of the block submatrices of A as follows:

$$A_{1,1} = \begin{bmatrix} \bar{A}_{0,0} & \bar{A}_{0,1} \\ \bar{A}_{1,0} & \bar{A}_{1,1} \end{bmatrix}, \quad A_{2p-1,2p-1} = \begin{bmatrix} \bar{A}_{2p-1,2p-1} & \bar{A}_{2p-1,2p} \\ \bar{A}_{2p,2p-1} & \bar{A}_{2p,2p} \end{bmatrix}$$

$$A_{2p-1,1}=\begin{bmatrix}\overline{A}_{2p-1,0} & \overline{A}_{2p-1,1}\\ \overline{A}_{2p,0} & \overline{A}_{2p,1}\end{bmatrix},\quad A_{1,2p-1}=\begin{bmatrix}\overline{A}_{0,2p-1} & \overline{A}_{0,2p}\\ \overline{A}_{1,2p-1} & \overline{A}_{1,2p}\end{bmatrix}$$

$$A_{1,i}=\begin{bmatrix}\overline{A}_{0,i}\\ \overline{A}_{1,i}\end{bmatrix},\quad A_{i,1}=\left(\overline{A}_{i,0},\overline{A}_{i,1}\right),\quad A_{2p-1,i}=\begin{bmatrix}\overline{A}_{2p-1,i}\\ \overline{A}_{2p,i}\end{bmatrix}$$

for $1<i<2p-1$

$$A_{i,2p-1}=\left(\overline{A}_{i,2p-1},\overline{A}_{i,2p}\right),\quad u_1=\begin{bmatrix}\overline{u}_0\\ \overline{u}_1\end{bmatrix},\quad u_{2p-1}=\begin{bmatrix}\overline{u}_{2p-1}\\ \overline{u}_{2p}\end{bmatrix},\quad 1<i<2p-1$$

After these rewriting of those sub-matrices along the boundary of matrix A, we can obtain the following theorem, which requires less zero sub-matrices in its assumption. The proof of this theorem is very similar to that of Teoremei 1.2.1.

Theorem 1.2.2 If $A_{i,2j}=0$ if $|i-2j|\geq 2$, $\overline{A}_{i,0}=0$ if $2\leq i\leq 2p$ and $\overline{A}_{i,2p}=0$ for $1\leq i\leq 2p-2$ then:

i) The iterative matrix of the AMS method satisfies:

$$\lambda\left(H^{-1}\tilde{B}\right)\subsetneq\lambda\left(\left(\ddot{D}+\omega\ddot{L}\right)^{-1}\left((1-\omega)\ddot{D}-\omega\ddot{U}\right)\right)$$

and:

$$\lambda\left(\left(\ddot{D}+\omega\ddot{L}\right)^{-1}\left((1-\omega)\ddot{D}-\omega\ddot{U}\right)\right)\setminus\{1-\omega\}\subsetneq\lambda\left(H^{-1}\tilde{B}\right)$$

ii) The iterative matrix of the AAS method satisfies:

$$\lambda\left(\tilde{B}^{-1}\left(\tilde{D}-\omega\tilde{A}\right)\right)\subsetneq\lambda\left(\ddot{B}^{-1}\left(\ddot{D}-\omega\ddot{A}\right)\right)$$

and:

$$\lambda\left(\ddot{B}^{-1}\left(\ddot{D}-\omega\ddot{A}\right)\right)\setminus\{1-\omega\}\subseteq\lambda\left(\tilde{B}^{-1}\left(\tilde{D}-\omega\tilde{A}\right)\right)$$

where:

$$\ddot{A}=\ddot{D}+\ddot{L}+\ddot{U}$$

$$\ddot{D}=\begin{bmatrix}\ddot{A}_{1,1} & & & \\ & \ddot{A}_{3,3} & & \\ & & \ddots & \\ & & & \ddot{A}_{2p-1,2p-1}\end{bmatrix}$$

$$\ddot{L}=\begin{bmatrix}0 & & & & \\ \ddot{A}_{3,1} & 0 & & & \\ \ddot{A}_{3,1} & \ddot{A}_{3,3} & 0 & & \\ \vdots & & \ddots & \ddots & \\ \ddot{A}_{2p-1,1} & & & \ddot{A}_{2p-1,2p-3} & 0\end{bmatrix}\qquad \ddot{U}=\begin{bmatrix}0 & \ddot{A}_{1,3} & \cdots & \ddot{A}_{1,2p-1}\\ & \ddots & \cdots & \vdots\\ & & \ddot{A}_{2p-5,2p-3} & \ddot{A}_{2p-5,2p-3}\\ & & 0 & \ddot{A}_{2p-3,2p-1}\\ & & & 0\end{bmatrix}$$

where, for $1 \le i \le p.$

$$\ddot{A}_{1,2i-1} = \overline{A}_{1,2i-1} - \overline{A}_{1,0}\overline{A}_{0,0}^{-1}\overline{A}_{0,2i-1} - \overline{A}_{1,2}\overline{A}_{2,2}^{-1}A_{2,2i-1}$$

$$\ddot{A}_{2p-1,2i-1} = \overline{A}_{2p-1,2i-1} - \overline{A}_{2p-1,2p}\overline{A}_{2p,2p}^{-1}\overline{A}_{2p,2i-1} - \overline{A}_{2p-1,2p-2}\overline{A}_{2p-2,2p-2}^{-1}A_{2p-2,2i-1}$$

and for $2 \le i \le p-1, m = 1 \quad or \quad 2p\text{-}1$

$$\ddot{A}_{2i-1,m} = \overline{A}_{2i-1,m} - A_{2i-1,2i-2}\overline{A}_{2i-2,2i-2}^{-1}\overline{A}_{2i-2,m} - A_{2i-1,2i}\overline{A}_{2i,2i}^{-1}A_{2i,m}$$

From Theorem 1.2.1 and 1.2.2, we established the eigenvalue relation between the matrix A and $\dot{A}$ or $\ddot{A}$. In order to obtain the convergence factors of the AMS method, we further discuss the relation between A and $\dot{A}$, as well as A and $\ddot{A}$ in following lemma.

Lemma 1.2.1 Suppose that:

$$A_{i,2j} = 0 \text{ and } A_{2j,i} = 0 \text{ for } |i-2j| \ge 2$$

then:

i) - if A is symmetric $\Rightarrow$ $\dot{A}$ is symmetric.

- if A is positive definite $\Rightarrow$ $\dot{A}$ is positive definite.

ii) assume $\overline{A}_{i,0} = 0, \quad \overline{A}_{0,i} = 0$ for $2 \le i \le 2p$ and $\overline{A}_{i,2p} = 0, \quad \overline{A}_{2p,i} = 0,$ for $1 \le i \le 2p-2.$ Then:

- if A is symmetric $\Rightarrow$ $\ddot{A}$ is symmetric
- if is positive definite $\Rightarrow$ $\ddot{A}$ is positive definite

iii) if A is an M – matrix $\Rightarrow$ $\dot{A}$ is an M matrix

iv) assume that $\overline{A}_{i,0} = 0, \quad \overline{A}_{0,i} = 0$ for $2 \le i \le 2p$ and $\overline{A}_{i,2p} = 0, \quad \overline{A}_{2p,i} = 0,$ for:

$1 \le i \le 2p-2.$ If A is an M – matrix $\Rightarrow \ddot{A}$ is an M – matrix.

Proof Since the proof of (ii) and (iv) is similar to that of (i) and (iii), we only prove (i) and (ii) here.

i) Because $A_{i,2j} = 0$ for $|i-2j| \ge 2$, it is obvious that $\dot{A}$ is symmetric.

From $\dot{u}^T\dot{A}\dot{u} = \sum_{i,j=1}^{p} \dot{u}_{2i-1}^T\dot{A}_{2i-1,2j-1}\dot{u}_{2j-1}$ we let:

$$u_{2i} = -A_{2i,2i}^{-1}\left(A_{2i,2i-1}\dot{u}_{2i-1} + A_{2i,2i+1}\dot{u}_{2i+1}\right)$$

and construct a vector u by putting these subvectors $\dot{u}_{2i-1}$ and u_{2i}. Then

$$\dot{u}^T\dot{A}\dot{u} = \sum_{i,j=1}^{p} \dot{u}_{2i-1}^T\dot{A}_{2i-1,2j-1}\dot{u}_{2j-1} = \sum_{i,j=1}^{2p-1} u_i^T A_{i,j}u_j$$

Hence, the positive definite of A implies the positive definite of $\dot{A}$.

iii) For any $\dot{b} = \left(\dot{b}_1^T, \dot{b}_3^T, \ldots, \dot{b}_{2p-1}^T\right)^T \ge 0$ there exists a vector $\dot{u}$ such that $\dot{A}\dot{u} = \dot{b}.$ Let u be a vector whose subvectors are defined from the vector of $\dot{u}$ and the solutions u_{2i} of:

$$A_{2i,2i}u_{2i} = -\left(A_{2i,2i-1}\dot{u}_{2i-1} + A_{2i,2i+1}\dot{u}_{2i+1}\right)$$

Then this vector u satisfies:

$$Au = b \text{ where } b = \left(b_1^T, 0, \dot{b}_3^T, 0, \ldots, 0, \dot{b}_{2p-1}^T\right).$$

Since $A^{-1} \geq 0$ and $b \geq 0$ it follows that $u = A^{-1}b \geq 0$. Thus $\dot{A}^{-1} \geq 0$. Note that $\dot{A}$ is positive definite. So the diagonal elements of $\dot{A}$ must be positive. Now we prove that the off-diagonal elements $\dot{A}$ are negative. Let $\dot{u}$ be the vector that only one component of $\dot{u}$ is one and the other components are zero. Let:

$$\dot{A}\dot{u} = \dot{b} \text{ where } \dot{b} = \left(b_1^T, 0, \dot{b}_3^T, 0, \ldots, 0, \dot{b}_{2p-1}^T\right) \text{ and}$$

$$A_{2i,2i}u_{2i} = -\left(A_{2i,2i-1}\dot{u}_{2i-1} + A_{2i,2i+1}\dot{u}_{2i+1}\right) \quad 1 \leq i \leq p-1$$

Because $A_{2i,2i-1}\dot{u}_{2i-1} + A_{2i,2i+1}\dot{u}_{2i+1} \leq 0$ and $A_{2i,2i}^{-1} \geq 0$ we have $u_{2i} \geq 0, 1 \leq i \leq p-1$. Then a vector u is defined and $Au = b$ with $b = \left(b_1^T, 0, \dot{b}_3^T, 0, \ldots, 0, \dot{b}_{2p-1}^T\right)$. From this equations, we can obtain that the component of $\dot{b}$ corresponding to the nonzero components of $\dot{u}$ must be strictly positive and other components of $\dot{b}$ are negative. Then $\dot{A}$ is an M – matrix.

Remarks We can use[22] the spectrum δ of the block Jacobi iterative matrices of $\dot{A}$ and $\ddot{A}$ to get the optimal:

$$\omega_{\text{opt}} = \frac{2}{1 + \sqrt{1 - \delta^2}}$$

This choice makes the convergence factor $\lambda = \omega - 1$ minimum. Since $\delta < 1$ we prefer to choose $1 < \omega < 2$ in AMS and AAS method in practice.

1.3 Iterative Subsolver for All Sub-problems

We write the matrix A as the sum of diagonal matrix D, lower triangular matrix L and upper triangular matrix U, $A = D + L + U = D + C$ where

$$D = \begin{bmatrix} D_{1,1} & & \\ & \ddots & \\ & & D_{2p-1,2p-1} \end{bmatrix} \qquad L = \begin{bmatrix} L_{1,1} & & & \\ A_{2,1} & & L_{2,2} & \\ \vdots & & \ddots & \ddots \\ A_{2p-1,1} & \cdots & A_{2p-1,2p-2} & L_{2p-1,2p-1} \end{bmatrix}$$

$$U = \begin{bmatrix} U_{1,1} & A_{1,2} & \cdots & A_{1,2p-1} \\ & U_{2,2} & & \vdots \\ & & \ddots \ddots & A_{2p-2,2p-1} \\ & & & U_{2p-1,2p-1} \end{bmatrix}$$

Then, the matrix $\tilde{A}$ has a corresponding decomposition denoted as,

$$\tilde{A} = \hat{D} + \hat{L} + \hat{U} = \hat{D} + \hat{C}$$

where:

$$\hat{D} = \begin{bmatrix} \hat{D}_{1,1} & & \\ & \ddots & \\ & & \hat{D}_{p,p} \end{bmatrix} \qquad L = \begin{bmatrix} \hat{L}_{1,1} & & & \\ \hat{A}_{2,1} & \hat{L}_{2,2} & & \\ \vdots & \ddots & \ddots & \\ \hat{A}_{p,1} & \dots & \hat{A}_{p,p-1} & \hat{L}_{p,p} \end{bmatrix}$$

$$\hat{U} = \begin{bmatrix} \hat{U}_{1,1} & \hat{A}_{1,2} & \dots & \hat{A}_{1,p} \\ & \hat{U}_{2,2} & & \vdots \\ & & \ddots \ddots & \hat{A}_{p-1,p} \\ & & & \hat{U}_{p,p} \end{bmatrix}$$

Here we let:

$$\hat{D}_{1,1} = \begin{bmatrix} D_{1,1} & \\ & D_{2,2} \end{bmatrix}, \qquad \hat{D}_{p,p} = \begin{bmatrix} D_{2p-2,2p-2} & \\ & D_{2p-1,2p-1} \end{bmatrix}$$

$$\hat{L}_{1,1} = \begin{bmatrix} L_{1,1} & \\ A_{2,1} & L_{2,2} \end{bmatrix}, \qquad \hat{L}_{p,p} = \begin{bmatrix} L_{2p-2,2p-2} & \\ A_{2p-1,2p-2} & L_{2p-1,2p-1} \end{bmatrix}$$

$$\hat{U}_{1} = \begin{bmatrix} U_{1,1} & A_{1,2} \\ & U_{2,2} \end{bmatrix}, \qquad \hat{U}_{p} = \begin{bmatrix} U_{2p-2,2p-2} & A_{2p-2,2p-1} \\ & U_{2p-1,2p-1} \end{bmatrix}$$

$$\hat{D}_{i} = \begin{bmatrix} D_{2i-2,2i-2} & & \\ & D_{2i-1,2i-1} & \\ & & D_{2i,2i} \end{bmatrix}, \quad \hat{L}_{i} = \begin{bmatrix} L_{2i-2,2i-2} & & \\ A_{2i-1,2i-2} & L_{2i-1,2i-1} & \\ A_{2i,2i-2} & A_{2i,2i-1} & L_{2i,2i} \end{bmatrix}$$

$$\hat{U}_{i} = \begin{bmatrix} U_{2i-2,2i-2} & A_{2i-2,2i-1} & A_{2i-2,2i} \\ & U_{2i-1,2i-1} & A_{2i-1,2i} \\ & & U_{2i,2i} \end{bmatrix}$$

Suppose that p subproblems are solved using point Jacobi iterative method with $\omega_i = \omega, i = 1,\dots,p.$ Then, the AAS algorithm can be written in the form:

$$\hat{D}\tilde{u}^{(k+1)} = (\hat{D} - \omega\hat{A})\tilde{u}^{(k)} + \omega f$$

and the AMS method can be represented by:

$$(\hat{D} + \omega\hat{L})\tilde{y}^{(k+1)} = (\hat{D} - \omega(\hat{D} + \hat{U}))\tilde{y}^{(k)} + \omega f$$

Theorem 1.3.1 Suppose that $D_{i,i}$ are invertible for $i = 1,\dots,2p-1$. Then we have:

$$\lambda\left(\hat{D}^{-1}(\hat{D} - \omega\hat{A})\right) = \lambda\left(D^{-1}(D - \omega A)\right) \cup \left(\bigcup_{i=1}^{p} \lambda\left(D_{2i}^{-1}(D_{2i} - \omega A_{2i,2i})\right)\right)$$

Proof Suppose that $\lambda\left(\hat{D}^{-1}(\hat{D}-\omega\hat{A})\right)$ and $\hat{u}$ is the corresponding eigenvector, i.e. $\lambda\hat{D}\hat{u}=(\hat{D}-\omega\hat{A})\hat{u}$.

If $u_{2i}\neq\hat{u}_{2i}$ for $1\leq i\leq p-1$, then we have $\lambda\in\lambda\left(D_{2i}^{-1}(D_{2i}-\omega A_{2i,2i})\right)$

If $u_{2i}=\hat{u}_{2i}$ for $i=1,\ldots,p-1$, then we have $\lambda\in\lambda\left(D^{-1}(D-\omega A)\right)$ and we already show that:

$$\lambda\left(\hat{D}^{-1}(\hat{D}-\omega\hat{A})\right)\subseteq\lambda\left(D^{-1}(D-\omega A)\right)\bigcup\left(\bigcup_{i=1}^{p}\lambda\left(D_{2i}^{-1}(D_{2i}-\omega A_{2i,2i})\right)\right)$$

Now we prove that:

$$\lambda\left(D^{-1}(D-\omega A)\right)\bigcup\left(\bigcup_{i=1}^{p}\lambda\left(D_{2i}^{-1}(D_{2i}-\omega A_{2i,2i})\right)\right)\subseteq\lambda\left(\hat{D}^{-1}(\hat{D}-\omega\hat{A})\right)$$

.Assume that:

$$\lambda\in\lambda\left(D^{-1}(D-\omega A)\right)\bigcup\left(\bigcup_{i=1}^{p}\lambda\left(D_{2i}^{-1}(D_{2i}-\omega A_{2i,2i})\right)\right)$$

If $\lambda\in\lambda\left(D^{-1}(D-\omega A)\right)$ and u is the associated eigenvector, i.e. $-\lambda Du+(D-\omega A)u=0$, then we construct an eigenvector $\hat{u}$ of $\hat{D}^{-1}(\hat{D}-\omega\hat{A})$, from the eigenvector u through letting $\hat{u}_{2i}=u_{2i}$ for $i=1,\ldots,p-1$.

If $\lambda\in\left(\bigcup_{i=1}^{p}\lambda\left(D_{2i}^{-1}(D_{2i}-\omega A_{2i,2i})\right)\right)$ and $\lambda\notin\lambda\left(D^{-1}(D-\omega A)\right)$, then an eigenvector $\hat{u}$ of matrix $\hat{D}^{-1}(\hat{D}-\omega\hat{A})$ is constructed by following procedure. Assume that v_{2i} is the solution of equation:

$$-\lambda D_{2i}v_{2i}+(D_{2i}-\omega A_{2i,2i})v_{2i}=0$$

i.e. v_{2i} is the eigenvector of $D_{2i}^{-1}(D_{2i}-\omega A_{2i,2i})$. By solving equation $-\lambda Du+(D-\omega A)u=w$ where $w=\left(0^T,\ldots,0^T,\omega v_{2i}^T A_{2i+1,2i},\ldots,\omega v_{2i}^T A_{2p-1,2i}^T\right)$ we obtain a vector u. Define $\hat{u}_{2i}=v_{2i}+u_{2i}$. The other subvectors of $\hat{u}$ are defined by $\hat{u}_{2j}=u_{2j},\ j=1,\ldots,i-1,i+1,\ldots,p-1$. Then, this $\hat{u}$ satisfies

$$-\lambda\hat{D}\hat{u}+(\hat{D}-\omega\hat{A})\hat{u}=0.$$

Therefore, $\lambda\in\lambda\left(\hat{D}^{-1}(\hat{D}-\omega\hat{A})\right)$.□

Theorem 1.3.2 Suppose that A is an M – matrix,

$$\left(D+\omega\begin{bmatrix}0 & & & \\ A_{2,1} & & & \\ \vdots & & & \\ A_{2p-1,1} & \cdots & A_{2p-1,2p-2} & 0\end{bmatrix}\right)^{-1}$$

exists and is nonnegative, and matrix:

$$D-\omega\begin{bmatrix} A_{1,1} & \cdots & A_{1,2p-1} \\ & \ddots & \vdots \\ & & A_{2p-1,2p-1}\end{bmatrix}$$

is nonnegative. Then $\hat{A}^{-1}$ and $(\hat{D}+\omega\hat{L})^{-1}$ exist and are nonnegative, and $\hat{D}-\omega(\hat{D}+\hat{U})$ is nonnegative. So the spectrum of the AMS iterative matrix AMS $(\hat{D}+\omega\hat{L})^{-1}(\hat{D}-\omega(\hat{D}-\hat{U}))$ is less than 1.

The proof of this theorem is similar to that of Lemma 1.2.1.□

Let all subproblems be solved by the SOR method with $\omega_i=1,\ i=1,\ldots,p.$

Then, the AAS method can be expressed by:

$$\left(\hat{D}+\omega\begin{bmatrix}\hat{L}_{1,1} & & \\ & \ddots & \\ & & \hat{L}_{p,p}\end{bmatrix}\right)u^{(k+1)}=\left((1-\omega)\hat{D}-\omega(\hat{L}+\hat{U})\right)u^{(k)}+\omega f\ .$$

Hence, the corresponding iterative matrix is:

$$\hat{J}=\left(\hat{D}+\omega\begin{bmatrix}\hat{L}_{1,1} & & \\ & \ddots & \\ & & \hat{L}_{p,p}\end{bmatrix}\right)^{-1}\left((1-\omega)\hat{D}-\omega(\hat{L}+\hat{U})\right).$$

The AMS method can also be written in the simple form:

$$(\hat{D}+\omega\hat{L})y^{(k+1)}=\left((1-\omega)\hat{D}-\omega\hat{U}\right)y^{(k)}+\omega f$$

with the iterative matrix:

$$\hat{S}=(\hat{D}+\omega\hat{L})^{-1}\left((1-\omega)\hat{D}-\omega\hat{U}\right)$$

Theorem 1.3.3 Assume that A is an M – matrix,

$$\left(D+\omega\begin{bmatrix}L_{1,1} & & & \\ & \cdot & & \\ & & \cdot & \\ & & & L_{2p-1,2p-1}\end{bmatrix}\right)^{-1}$$

exists and is nonnegative, and:

$$(1-\omega)D-\omega\begin{bmatrix}0 & & & \\ A_{2,1} & & & \\ \vdots & & & \\ A_{2p-1,1} & \cdots & A_{2p-1,2p-2} & 0\end{bmatrix}-\omega U$$

is nonnegative. Then:

$$\hat{A}^{-1} \text{ and } \left(\hat{D}+\omega\begin{bmatrix}\hat{L}_{1,1} & & \\ & \ddots & \\ & & \hat{L}_{p,p}\end{bmatrix}\right)^{-1}$$

exist and are nonnegative. So the spectrum of the AAS iterative matrix $\hat{J}$ is strictly less than 1.

The proof is similar to that of Lemma 1.2.1.□

Denote:

$$M=(D+\omega L)^{-1}((1-\omega)D-\omega U)$$

and:

$$M_{2i,2i}=(D_{2i,2i}+\omega L_{2i,2i})^{-1}((1-\omega)D_{2i,2i}-\omega U_{2i,2i}).$$

Theorem 1.3.4 Assume that all $D_{i,i}$ are nonsingular. Then, we have:

$$\lambda(\hat{S})=\lambda(M)\cup\left(\bigcup_{i=1}^{p-1}\lambda(M_{2i,2i})\right)$$

Proof Assume that $\lambda\in\lambda(\hat{S})$ and $\hat{u}$ be the corresponding eigenvector, i.e.

$$\lambda\hat{u}=\hat{S}\hat{u},\quad \lambda(\hat{D}+\omega\hat{L})\hat{u}=((1-\omega)\hat{D}-\omega\hat{U})\hat{u}$$

If $u_{2i}\neq\hat{u}_{2i}$ for some $1\le i\le p-1$ then $\lambda\in\lambda(M_{2i,2i})$. If $u_{2i}=\hat{u}_{2i}$ for all $1\le i\le p-1$ then $\lambda\in\lambda(M)$. Hence, $\lambda(\hat{S})\subseteq\lambda(M)\cup\left(\bigcup_{i=1}^{p-1}\lambda(M_{2i,2i})\right)$.

Now we show that:

$$\lambda(M)\cup\left(\bigcup_{i=1}^{p-1}\lambda(M_{2i,2i})\right)\subseteq\lambda(\hat{S}).$$

Let $\lambda\in\lambda(M)$ and u be the associated eigenvector, i.e.

$$\lambda u=Mu \text{ and } \lambda(D+\omega L)u=((1-\omega)D-\omega U)u$$

Define $\hat{u}$ by letting $\hat{u}_{2i}=u_{2i}$. Then, this $\hat{u}$ is the eigenvector of $\hat{S}$ and $\lambda\hat{u}=\hat{S}\hat{u}$. Thus, $\lambda(M)\subseteq\lambda(\hat{S})$.

Assume $\lambda \in \lambda(M_{2i,2i})$, $\lambda \notin \lambda(M)$. Denote v_{2i} to be the corresponding eigenvector, i.e. $\lambda v_{2i} = M_{2i,2i}$. We solve the following equation and get a solution vector u from:

$$(\lambda(D+\omega L)+\omega U+(\omega-1)D)u = w$$

where $w = -\lambda\omega(0^T,\ldots,0^T,v_{2i}^T A_{2i+1,2i},\ldots,v_{2i}A_{2p-1,2i})^T$. Since $\lambda \notin \lambda(M)$, this problem has only one solution. We define a new vector by letting:

$$\hat{u}_{2j} = \begin{cases} v_{2j}+u_{2i} & j=i \\ u_{2j} & j \neq i \end{cases}$$

This $\hat{u}$ satisfies $\lambda\hat{u} = \hat{S}\hat{u}$. Hence, $\tilde{u}$ is the eigenvector of $\hat{S}$.

From above theorems and lemmas, we conclude that the convergence factors of the AAS method and AMS method are almost the same as the block Jacobi method and the SOR method.

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VALUE-ADDED SERVICES OF LOGISTICS CENTERS IN PORT AREAS

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Abstract: *Generally, the function of a port as a node in the transport chain depends on its location and on the economic and technical developments that exist in its hinterland. Modern production techniques and consumption patterns increase the use of transportation systems beyond levels suggested purely by the growth in trade and commerce. As a result, more specialized handling, storage and other logistics facilities are needed. This process of specialization and adjusting to customer's demands, which has taken place over the last two decades in most Western countries, is now taking place with even greater speed in the emerging economies.*
More and more, the ports are becoming an important part of the so-called integrated logistics chains. In order to stay on the market, the ports have to diversify their activities and services. Instead of offering traditional services, they have to focus more on implementing new value-adding logistics services. Port's managers need to work together with local authorities, law makers, manufacturers, investors in order to create a favorable environment to put all this in practice. Value-added logistics services in port areas will play a key role in the future in bringing more revenue to specific hinterlands.

Key words: *value-added service, cargo, port, ship, logistic center*

Value-added logistics services

In the 1970s, almost every port provided the same basic package of services to almost every customer. The late 1980s saw the emergence of major changes. Customers began to ask ports to provide a greater variety of services. Providing value-added services is a powerful way for ports to build a sustainable competitive advantage. Shippers and port customers are becoming increasingly demanding and now they tend to look at value-added logistics services as an integral part of their supply chain. In the recent years, many important international shipping companies have merged into bigger and stronger entities. These newly emerged entities are nowadays very important players on the transportation market and their needs have increased accordingly. As a result, ports must attempt to satisfy these needs by offering differentiated services. This poses a particular challenge for port management.

These days, the commercial success of a port could rise from a productivity advantage in traditional cargo-handling service, from value-added services, or from a combination of the two. The most productive ports will be those that are equipped to handle large cargo volumes and/or significantly reduce unit costs through efficient management. Shippers and carriers select individual ports not only based on their cargo handling service capabilities, but also on the benefits they are capable of "delivering". Unless a port can deliver benefits that are superior to those provided by its competitors in a functional aspect, port customers are likely to select ports based merely on price. This fact raises the question of how a port can achieve value differentiation.

Various studies show that the most successful ports are those that not only have a productivity advantage in cargo-handling services, but that also offer value-added services. Even though, it continues to be a need for ports that provide the basic, traditional cargo-handling function, and that there are still many customers for such services. Perhaps it is for this reason that many ports in developing countries still concentrate on improving their productivity with regard to traditional port functions, instead of building up value-adding logistics services. However, it is clear that, in the future, there will be fewer ports that prosper only in this area. Rather, we will see the dominance of superior service leaders that possess both a productivity advantage and a value-added service advantage. In between traditional service ports and superior service ports are the leading-edge service ports. These are the ports that are on their way to becoming superior service ports.

A number of ports have responded to this trend by focusing on value-added services as a mean of gaining a competitive edge. In this content, value-added service refers to the process of developing relationships with customers through the provision of a customized offer, which may include many aspects of value-added activities.

It is very advantageous for a port to be as well a logistics centre, since the logistics centers can attract cargo that can be shipped through the port. There is a direct positive correlation between cargo flows at the logistics centers and the number of ships calling at the port. In other words, the cargo attracts the ships, and the ships attract the cargo. The port benefits by generating increased revenue and creating additional jobs. The port can profit not only from the logistics centre itself, but also from the increased flow of cargo through the port. Thus, an ideal port should provide a diverse range of services that are highly integrated. As such, there is a need to seriously consider the increasing importance of ports in logistics management.

From the port's point of view, creating new services boosts the port's economic performance as well as its attractiveness to existing and potential clients. This, in turn, can help maintain and improve a port's competitive position. When assessing port's services, it is important to pay extra attention to the value adding potential of the services. This potential can vary product by product and activity by activity. Numerous activities can be classified as "Value Added Services"(figure 1).

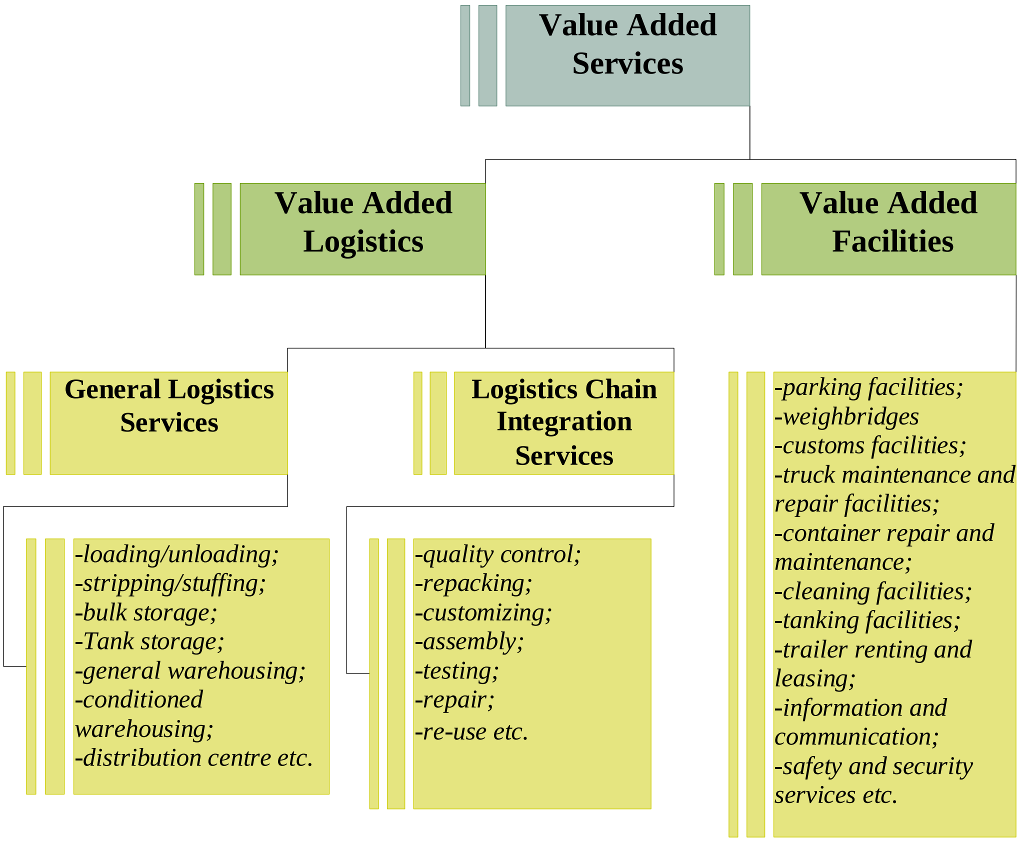

Figure 1 Overview of Value Added Services in Ports

Value Added Services can be divided into *Value Added Logistics* (VAL) and *Value Added Facilities* (VAF). Value Added Logistics has two major components: General Logistics Services (GLS) and Logistics Chain Integration Services (LCIS).

General Logistics Services are, among other activities, loading/unloading, stuffing and stripping, storage, warehousing and distribution. These are the more traditional logistics activities, and do not directly affect the nature of the product as it moves through the port. The General Logistics Services are provided nowadays by most of the ports all around the world. Beyond these traditional activities, more complex Logistics Chain Integration Services are being developed. To carry out activities that manufacturers do not consider part of their core business, logistics service providers may take over parts of the production chain (e.g., assembly, quality control, customizing and packing) and after sales services (e.g., repair and re-use). However, LCIS are only appropriate for certain types of goods. The products that have the highest potential to benefit from such services include: consumer electronics, pharmaceuticals, chemical products (except for those carried in bulk), clothing, cosmetics and personal care products, food, machinery and control engineering products. In Western Europe, Rotterdam and Hamburg are the leading ports in providing Logistics Chain Integration Services.

The second group of Value Added Services - Value Added Facilities (VAF) - is a very diverse group. These types of activities cannot generally be assigned to a particular type of product or freight flow. It is possible, however, to impute a certain "VAF-potential" by analyzing freight flows such as dry and liquid bulk, general cargo, containerized cargo and roll-on/roll-off. The existence of a large container terminal might create the economic basis for establishing container repair facilities. The container ships are big consumers of heavy fuel oil, diesel oil and lube oil (e.g., Maersk Line, the world's biggest container carrier line, is the world's second largest consumer of fuel oil - spending about 6 billion USD a year only on fuel), therefore they generate lots of quantities of sludge. In this respect, sludge collecting facilities can be established. Handling vast quantities of chemicals requires port reception facilities; substantial roll-on/roll-off traffic might justify truck maintenance and repair shops. Bunkering facilities for ships, if provided, can make as well a serious difference.

Usually, containerized and general cargos typically have the highest VAL potential. General Logistics Services and the Logistics Chain Integration Services have the best opportunity to serve these cargos. The VAL potential for roll-on/ roll-off is very limited. Trucks with drivers are too expensive to be delayed while the cargo is modified; additionally, these loads are usually customer-tailored. Value Added Facilities, such as tanking, cleaning, repair, parking, security, renting and leasing facilities have a better potential to serve the roll-on/roll-off market. Dry and liquid bulk flows have the lowest potential for both VAL and VAF (figure 2).

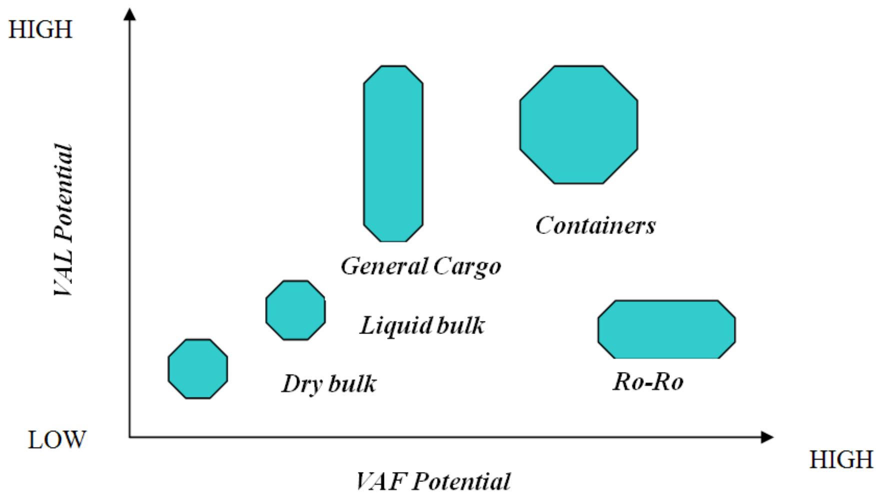

Figure 2 Potential for VAL and VAF

Conclusions

In the last decades, the market has changed drastically. The newly reformed and customer orientated ports have been real catalysts for the globalization process. Nowadays, however, it is more difficult for ports to compete on the basis of cargo-handling services. Modern technologies implemented in port's day-by-day activities have radically improved the quality of the services. But even though this new technology may provide a window of opportunity for productivity improvement, in many cases that same technology is also available to competitors. It is no longer possible to compete effectively on the basis of basic, traditional functions. Therefore, there is a need for ports to seek out new means of gaining a competitive edge and the most accessible way is to implement new value-adding logistics services.

As it has shown above, there is a multitude of value-adding logistics services to choose from. The real challenge is to determine which ones are to be implemented and how all this are going to be put in practice. This is the part where the port managers are playing the key role. It is their duty to bring to the table all the parties involved and to make the most out of the port's possessions and characteristics in order to improve its activity.

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CARDIORESPIRATORY FITNESS ASSESSMENT

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Abstract: *The purpose of this study was to identify and explain some commonly used methods to asses' cardio respiratory fitness. Additionally, the assessment must have minimal equipment and be moderately accurate. It was also shown that the field tests to asses CR fitness are based on scientific data that addresses such areas as relative and absolute measures of O^2, ventilatory break point, or onset of blood lactic acid, and work rates. And finally, the various methods in the Naval Academy uses to assess CR fitness were presented.*

Key wards: *cardio respiratory fitness, methods of assessing the level of CR, work rate, VO_2*

PURPOSE: To identify and explain common methods of assessing cardiorespiratory fitness which can be performed with little equipment and in a relativity short period of time.

TRAINING OBJECTIVES:

A. Identify the best measure of cardiorespiratory endurance, define VO_2 in relative and absolute terms, and calculate the VO_2 given the necessary data.

B. Define the ventilatory break point.

C. Identify factors important in administration, recording, and interpreting maximal and submaximal cardiorespiratory.

D. Present alternate in cardiorespiratory assessment events, such as bicycle ergometer, 2.5 – mile walk and the 800 – yard swim.

I. Max VO_2

As discussed in the class on the oxygen transport system, the single, best measure of a person's level of cardiorespiratory (CR) fitness is the maximum ability to consume oxygen (max VO_2) during very intense exercise. It is the common measure that allows comparing the results compiled using different assessment methods. A submaximal test given to a soldier can be used to predict with fair accuracy (+ or – 10%) the status of the oxygen transport system, or aerobic system.

II. VO2 and max VO2 Expressed in Relative and Absolute Terms

A. VO2 and max VO2 are expressed as liters of oxygen consumed per minute (an absolute measure) or as the average amount of oxygen consumed by each kilogram of body mass every minute, a relative measure expressed in milliliters of oxygen per kilogram of body weight per minute ($m10_2$/kg/min), respectively.

B. These are provided as a means of classifying a soldier's level of CR fitness based on actual on estimated max VO_2. it is noted that estimated values are not as accurate (fair accuracy + or – 10%) as values obtained by the direct measurement of max VO_2. Also note that the value are expressed in the relative terms, i.e., ml o_2/kg x min.

III. The calculation of VO_2

A. To determine oxygen uptake (VO_2) for a given time during exercise or rest, substract the volume of oxygen expired during that time from the volume of oxygen inspired during that knowledge of the volume of air inspired and expired during that period and the concentration of oxygen in the inspired and expired air, as shown in the following formulas.

VO_2 = (volume air inspired x concentration o_2, in inspired air) – (volume air expired x concentration O_2 in expired air)

B. The concentration of O_2 in inspired air is 0.2093 or, expressed as a percentage, 20.93%. The concentration of O_2 in the expired air is measured with an oxygen analyzer. The volume of expired and inspired air during a given time period can be measured by one of several methods.

C. An example of how to determine the VO_2 using the above equation is provided using the values of 100 liters of inspired and expired air, a value of 16.00% for the concentration of oxygen in the expired air, and a time period for the measurements of the minute.

VO_2 = (100 liters inspired x 20.93% O_2) – (100 liters expired x 16.00% O_2)
Convert the percentages to fractions before multiplying:

VO_2 = (100 liters inspired x 0.2093) – (100 liters expired x.1600)
After multiplying, the following is obtained:
VO_2 = (20.93 liters of O_2inspired) – (16.00 liters of O_2 expired)

After subtracting the expired oxygen from the inspired oxygen, the VO_2 (the volume of O_2 used is 4.93 liters of oxygen during the one minute measured).

IV. Methods of assessing the level of cardiorespiratory fitness

The routine method for assessing an individual's level of CR fitness is to test the individual on a treadmill, the 2-mile run, or by such alternate tests the individual on a treadmill, the 2 – mile run, or by such alternate tests on the bicycle, 2.5 – mile walk tests, or the 800 – yard swim. Some tests are classified as maximal tests and some, which do not require an all-out effort by the individual, are classified as submaximal tests. Several tests are used in the army to assess a soldier's level of CR fitness.

A. Treadmill. There are many treadmill tests used to assess CR fitness. Treadmill testing increases the workload as the test proceeds. The increasing intensity is accomplished by increasing the speed of the treadmill and/or by increasing the slope of treadmill. The initial speed and slope are within an individual's basic walking capacity.

To determine CR fitness, the workload is made more difficult until the runner reaches a point where continuing the test is beyond his/her capacity. Information is gathered and calculations are made to determine max VO_2.

B. 2-Mile-run times and maxVO_2. A number of performance tests have been devised for testing large groups of individuals in fields situation. These tests are practical, inexpensive and easy to administer. The results are quite accurate when the subjects are properly motivated, as these are CR tests and require maximum efforts by the student. The basis for using these tests is the strong correlation that exists between the max VO_2 determined in the laboratory and 2-mile-run times.

The 2-mile run event tests cardiorespiratory (aerobic) endurance and the muscular endurance of the muscles. It requires the students to complete the run without any physical assistance. At the start, all students line up the starting line. On the command "Go", the clock starts and the student begins running at his/her own place. The student completes the 2-mile course in the shortest time possible. Although walking is authorized, it is strongly discouraged. If a soldier is physically assisted in any way, (i.e., pulled, pushed, picked up and/or carried) or levels the designated running course for any reason, he/she is disqualified. It is acceptable to pace another student during the 2-mile run. However, no physical contact with the paced student is allowed. Running ahead of, along side of, or behind the tested student, while serving as a pacer is permitted as long as there is no hindrance to other student taking the test. Cheering or reading out elapsed time is also permitted. After completing the run, the student returns to the designature area for the cool – down and stretch.

As mentioned previously, there exists strong correlation between the max VO_2 determined in the laboratory and 2-mile-run times. Below are equations that can be used to determine a student's max VO_2 based in the 2-mile-run time:

MALES: max VO_2 = 99.7 – [3.35 x time]
FEMALES: max VO_2 = 72.9 – [1.77 x times]

Time = the 2-mile run time in minutes, expressed as a decimal

To calculate a run time expressed as a decimal, divide the seconds in the run time by 60 seconds in 1 minute. For example, a 2 mile run time of 14 minute, 30 seconds is expressed as 14.5 minutes (30/60 = 0,25) of 11 minutes and 15 seconds is expressed as 11.25(15/60 = 0,25)

Exemple:

Male max VO_2 = 99,7 - (3,35 x2 mile run time) = 99,7 – (3,35 x 11,25)
= 99,7 – 37,69 = 62,01 ml O_2/ kg/min.

Femele max VO_2 = 72,9 – (1,77 x 2 mile run time) = 72,9 – (1,77 x 14,5)
= 72,9 – 25,66
= 47,24 ml O_2/ kg/min.

C. The 800-yard swim test. This test alternate aerobic test is used for students who are on profile and can not run. The test is used because there is a correlation between the aerobic events.

Instruction: The instructor reads the following statement: the 800-yard swim is used to asses the level of aerobic fitness. You will begin in the water, no diving is allowed. The body must be in contact with the wall of the pool. On the command "start" the clock will start and you begin swimming, using any stroke or combination of stroke desired. Any type of turn is authorized. The score is less than, that listed for age. Walking on the bottom to recuperate is authorized.

Timing Techniques: the event supervisor also serves as timer. The commands “Get set” and “Go” are used. Two stopwatches are used in case one fails. As the students near the finish, the event supervisor begins calling off the elapsed time in minutes and seconds, as in the following example: “nineteen-eleven, nineteen-twelve, nineteen-thirteen”, and so on. The time is recorded when each student touches the end of the pool on the final lap or when the student crosses a line established as the 800-yard mark.

Scorers Duties: the scorers observe the swimmers assigned to them. They must be sure that each swimmer touches the bulkhead on every turn. The scorers record each student’s time in the block provided for the 2-mile run on the scorecard. A notation is made in the comment block to identify the time as an 800-yard-swim time. If the pool length is measured in meters the exact distance is converted to yards. To convert meters to yards, multiply the number of meters by 36, i.e., (meters x 39.37)/36 = 437.4 yards.

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METHODS IN THE ROMANIAN NAVAL PENTATHLON PERFORMANCE FOR UTILITY SWIMMING

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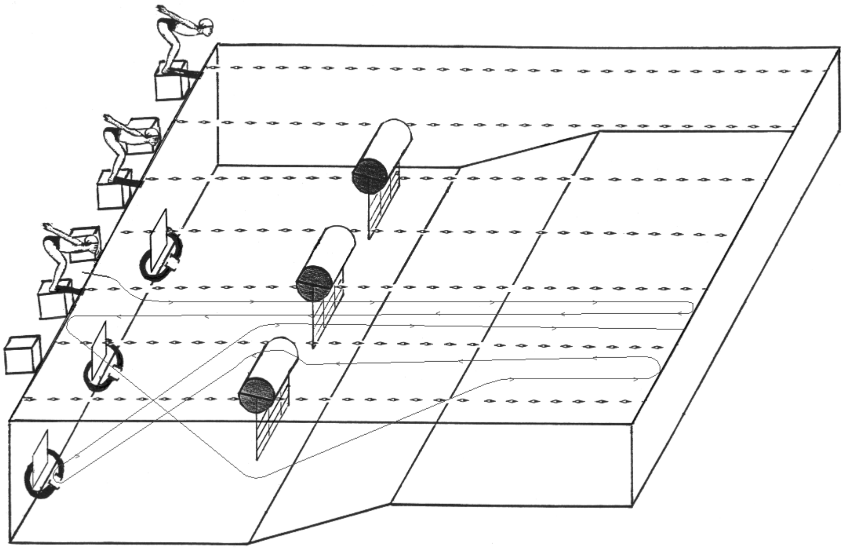
Abstract: *In this paper we try to present a model of training for the naval pentathlon. Utility swimming race it's the most important event in the naval pentathlon. The increase results for this event it's very important in ergonomic the total points to the naval pentathlon.*

Keywords: *Lifesaving race, seamanship race, utility swimming race, amphibious cross-country race.*

Introduction

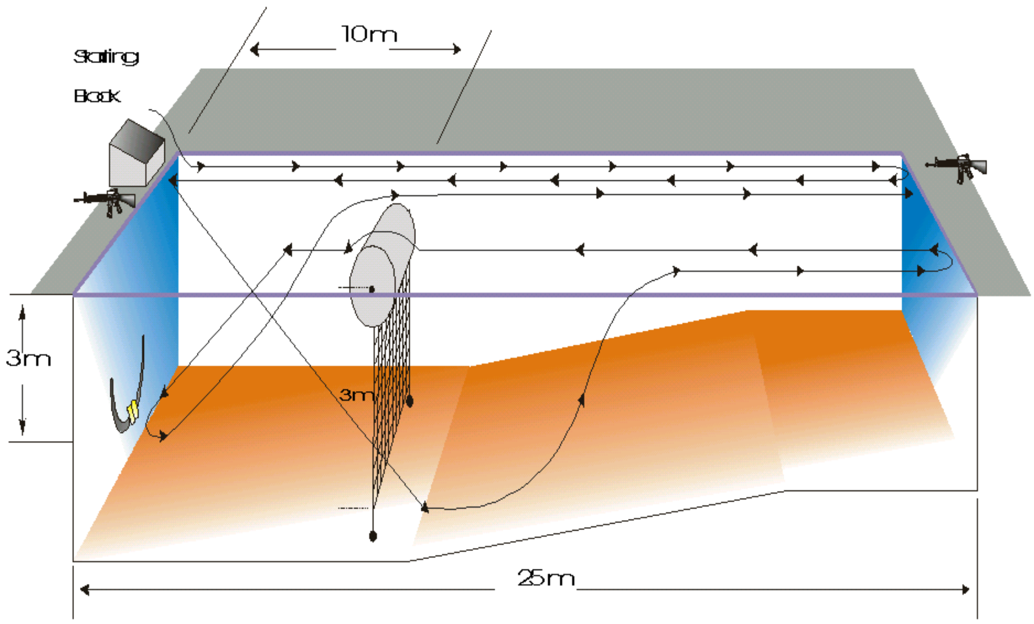
Naval Pentathlon is an individual, male and female, competition consisting of the following five events: obstacle race; lifesaving swimming race; utility swimming race; seamanship race; amphibious cross-country race. An "Individual champion" is determined by the overall result in the five events. The team champion is determined by adding the individual results of a country's team. The regulations prescribe the way a CISM (International Council of Military Sports) military world championship shall be conducted. I'm present in the next principal characters about 3rd event in naval pentathlon: u*tility swimming race*. This race is conducted in the same pool as the one used for the Life-Saving Race. The overall distance of the race is 125/100m, during which the competitor shall perform six/five (6/5) separate features (picture 1). Utility swimming race included 8 features (Naval Pentathlon Regulation, 2003) [1]:



Picture 1 Utility swimming race area

<u>**Feature No. 1**</u> **- Start.** Dive from the starting block/edge of the pool.
<u>*Condition*</u> - Start and swim 25 m, free style.
<u>**Feature No. 2**</u> **- Rifle carrying.** Distance from the start – 25 m. A rifle with a weight of about 3kg. It shall be laid at the edge of the pool, in a fix way (see sketch) by the organiser.
<u>*Condition*</u> - The competitor shall take the rifle from the edge of the pool (25 m) and carry it unaided until he reaches the opposite side of the pool (50 m), where he has to leave the rifle outside the pool <u>after</u> touching the wall. If the competitor fails to leave the rifle outside the pool or leaves it before first touching the wall there will be a 15 sec. penalty for each fault.
<u>**Feature No. 3**</u> **- Passing under a net obstacle.** Distance from the start - 60/35 m. A net with its lower edge at least 70cm from the bottom. With a chain reaching from each side down to the bottom.
<u>*Condition*</u> - Pass under the net. If the competitor fails to pass under the net he must, without disturbing any other competitor, swim at the side of the net (preferable on the left lane). If the competitor fails to pass under the net there will be a 15 sec. penalty.
<u>**Feature No. 4**</u> **- Pass over the barrel obstacle.** Distance from the start - 90/65 m. A cylindrical float made of wood or metal, as shown in sketch, held in position by lane lines. The cylinder shall be covered with neoprene or carpet and freely rotate and submerge when a competitor passes over it. Height above the water is 25 cm.
<u>*Condition*</u> - Pass over the barrel. If the competitor fails to pass over the barrel he must swim, at the side of the barrel, on the right side lane. If the competitor fails to pass over the barrel will be a 15 sec. penalty.
<u>**Feature No. 5**</u> **- Underwater work.** Distance from the start - 100/75 m. Two hoses at the end of the pool. The joined ends of the hoses are 3.00 m from the surface. Should the pool be only less than 3.00 m deep, the hoses shall be placed at the bottom. The hoses should be attached to the pool or weighted 50 cm from the ends to limit upward movement. The two couplings shall be of the fixed joint type.

The two hoses shall be standard shipboard fire-fighting hoses. The sponsoring nation shall provide participating nations with the specifications of the type of couplings to be used.
Condition - Uncouple the hoses. If the competitor fails to uncouple the hoses there will be a 15 sec. penalty.
Feature No. 6 - **Final sprint.** Distance from the start 100-75 m. Distance 25 m.
Condition - Swimming free-style. The race is finished when the swimmer touches the end of the pool (picture 2).



Picture 2 Obstacles in Utility swimming race area

Methods

For naval pentathlon, specific preparation and competition models were allocated to all events. From thee ideas, we propose a scientific program allocated in macro cycle. This macro cycle included five mezzo cycles:

Introductive mezzo cycles: 01 October – 29 October - 4 weeks;
Remaking introductive mezzo cycles: 30 October – 11 December - 6 weeks;
Base mezzo cycles: 12 December – 12 March - 14 weeks;
Below contest mezzo cycles: 13 March – 3 April – 3 weeks;
Contest mezzo cycles: 4 April – 13 April - 2 weeks.
The structure this macro cycles present in table nr.1.

Table nr 1
The training model for Romanian naval pentathlon team in macro cycles period

Mezzo cycles		**Introductive mezzo cycle**	**Remaking introductive mezzo cycle**	**Base mezzo cycle**	**Below contest mezzo cycle**	**Contest mezzo cycle**	**Total**
Weeks		4	6	14	3	2	**29**
Nr. training /week.		8	8	10	8	5	
Total training/mezzo cycle (hour)		32 (64)	48 (96)	140 (280)	24 (48)	10 (20)	**254 (508**
Obstacle race	Nr. hour total obstacle race (weeks).	4	3	5	4	2	
	Total hour/mezzo cycle	**16**	**18**	**70**	**12**	**4**	**120**
Lifesaving swimming race	Nr. hour total lifesaving race (weeks).	3	4	4	3	2	
	Total hour/mezzo cycle	**12**	**24**	**56**	**9**	**4**	**105**
Utility swimming race	Nr. hour total utility swimming race (weeks).	3	2	3	4	1	
	Total hour/ mezzo cycle	**12**	**12**	**42**	**12**	**2**	**80**
Seamanship race	Nr. hour total seamanship race (weeks).	4	4	5	4	2	
	Total hour/mezzo cycle	**16**	**24**	**70**	**12**	**4**	**126**
Amphibious cross-country race	Nr. hour total amphibious cross-country race (weeks).	3	4	2	3	2	
	Total hour/mezzo cycle	**12**	**24**	**28**	**9**	**4**	**77**

Results

In the first part of macro cycles we test witness group and experiment group at all the moments that include seamanship race. To the finish of 29 weeks, we realize an increase of average for experiment group comparative with witness group. All this is presented in next table. The model of macro cycles was adapted by W.Maglissho, 2003.

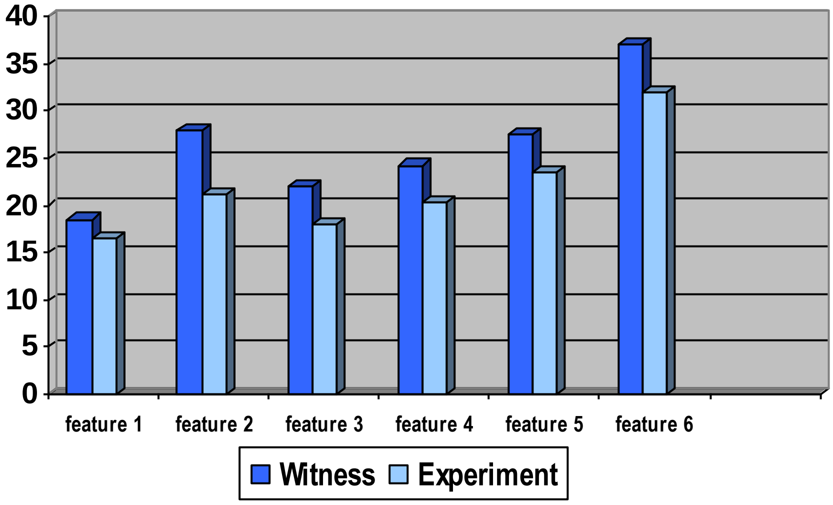
Table nr. 2
Comparative analyses in average characteristic utility swimming race event

Initial testing			
Nr.	**Parametric**	$\overline{X} \pm m$	
		Witness group	**Experiment group**
1	Start. (sec.)	19,50 ± 0,53	17,80 ± 0,22
2	Rifle carrying. (sec.)	29,60 ± 1,19	25,20 ± 0,88
3	Passing under a net obstacle. (sec.)	23,40 ± 1,05	22,00 ± 1,08
4	Pass over the barrel obstacle. (sec.)	28,25 ± 2,01	26,25 ± 0,55
5	Underwater work. (sec.)	29,00 ± 1,15	28,50 ± 0,35
6	Final sprint. (sec.)	41,00 ± 1,76	38,00 ± 0,86

Final testing			
Nr.	**Parametric**	$\overline{X} \pm m$	
		Witness group	**Experiment group**
1	Start. (sec.)	18,50 ± 1,53	16,50 ± 0,53
2	Rifle carrying. (sec.)	28,00 ± 1,89	21,20 ± 2,69
3	Passing under a net obstacle. (sec.)	22,00 ± 2,08	18,00 ± 3,08
4	Pass over the barrel obstacle. (sec.)	24,25 ± 2,03	20,25 ± 1,03
5	Underwater work. (sec.)	27,50 ± 0,15	23,50 ± 0,85
6	Final sprint. (sec.)	37,00 ± 2,12	32,00 ± 1,12

Discussion and conclusions

Increase the volume for utility swimming race but not only demonstrate that methods were realist implemented in training lesson. We present it in next graphic.



Graphic nr 1 Dynamics results features in utility swimming race

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THE EVOLUTION OF NAVY EDUCATION DURING THE INTER-WAR PERIOD

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Abstract: *Given the opportunity of the 136 years celebration of Romanian Navy Education, we would like to go back in time to depict the context, the necessity and the contribution of the ones involved – the state's institutions: the king, the parliament, the army and so on – in this commune effort regarding the progress of civilization and the country cultural development.*

Keywords: navy, education, inter-war period

1 The Transition Towards a Romanian Navy Education

Following the Independence War and the Berlin Congress (1878), in the context of a modernizing process for Romania that involves the growing role of military and commercial fleet as a power factor helping the protection and promoting economical, geopolitical and security interests belonging to riverside states, we too made the first steps towards the identification and institutionalization of adequate means of forming navy people, towards developing and perfecting a specialized navy education to satisfy society's practical exigencies.

Until a national education institution for the navy had been established, the army and military fleet leaders had opted for temporary solutions which involved sending the future navy cadets to attend famous foreign naval academies classes or learning directly onboard of military ships belonging to states with well-established traditions in this field: France, Italy, Germany, England, Russia and Austria-Hungary.

The instructions of the War Ministry (1900), regarded as historical documents, specify that the goal of this practice was that of "maintaining our instruction at the same level with the one abroad" [1], and "preparing officers to become professors and specialists" in order to push the military Romanian education at European standards. [2]

The professional performances and the military conduit of the ones sent abroad were closely watched by Romanian authorities. Due to the control rules for the officers sent to studies in foreign countries (18/12/1880), the diplomatic corp was supposed to maintain connection with the naval academies in question and report at least monthly about their mission. The cadets having weak results or inappropriate conduit were asked to come back and billed for the transport expenses which were at the same level with a captain indemnization – 300 lei. [3]

The selection and admission exam for these kinds of studies were held with great responsibility and involved oral and written exams under the supervision of a committee of Great Major State level. [4]

This way, during 1859 and 1877, there were 268 people sent abroad (140 in France, 29 in Italy, 25 in Belgium, 21 in Prussia, 4 in Austria-Hungary, 1 in Russia, 1 in Egypt) and between 1878 and 1914 only 98 officers of which just 16 for the navy.

Historical documents – especialy the anual ones belonging to the interbelic military, name many people that were specialised with the help of Military Navy: Ion Murgescu, Dumitru Poenaru, Ion Siropol, Constantin Bălescu, Constantin Perițeanu, Paul Rădulescu, Nicolae Mardare, Gheorghe Koslinski, Mihail Constantinescu, Constantin Mănescu, Nicolae Dumitrescu-Maican, Alexandru Martac (1865-1908, The Navy School of Brest); Mihai Mihăilescu, Nicolae Negru, Constantin Niculescu-Rizea, Aurel Cătuneanu, Vasile Scodrea, August Roman, Nicolae Păun, Gheorghe Dumitrescu (1884-1914, The Navy School of *Livorno*); Romulus Ştefănescu, Radu Irimescu, Virgiliu Popescu, Nicolae Steriopol, Alexandru Dumitrescu, Paul Zlatian, Alexandru Bardescu (1908-1914, *The Navy School of Kiel*); Ion Camenită, Ion Popovici, Ion Gheorghe (1897-1905, *The Mechanics Navy School of Toulon*); George Chirovici (The Trade and Nautic Academy *of Triest*); Nicolae Dumitrescu-Maican, Vasile Urseanu, Constantin Ciuchi, Petre Bărbuneanu, Nicolae Negrescu, Emanoil Koslinski, Ionescu Johnson Virgil Bucur (aboard the french, italian, english and russian ships). [5]

2 The Fleet School – the first Romanian navy education institution

If until the end of the Independence War, the preparation and specialization were basically solved through the foreign naval academies and fleets, after this occurrence, almost the entire training process was held by national military education institutions.

This way, through the Decision no.11/17 taken in November 1872, the authorities demanded the city of Galati to host the first military navy education institution – The Fleet School. [6]

This school, with personnel of highly experienced marine officers, represented a two years stage for officers and non-commissioned officers wannabes. The principle was the fleet's commander. The people in attendance were selected upon high standards and had their preparation schedule divided in two distinct but complementary periods: a) the theoretical stage (15th of November – 1st of April); b) the practical stage (15th of April – 11th of September, aboard military ships). The theoretical preparation included field related special disciplines, but also general culture classes: trigonometry, linear drawing, notions about fire holes or the sails installment, about steam powered machines and the international signals code, as well as foreign languages like French, Italian and German. The practical training was onboard, during instruction marches.

3 The Naval Academy

After a decade, the Fleet School was closed, and in 1881 - through the decision no.2408 from October - at general's George Slăniceanu initiative, the Navy Children School was born.

Even the school's stage was 3 years in length and the facility had a capacity of 20 pupils, it become functional only on the 31^{st} of January, next year.

At the same performance level with the one found in the military education institutions of Europe's more developed countries, our own education plan consisted of: a) general disciplines (Romanian language, math's, history, geography, cosmography, physics and chemistry notions, geometry and trigonometry, foreign useful languages and so on); b) disciplines specific to the institution's naval profile (designations, theories and applications with different guns and sails; the international Morse code, mechanical engineering, body exercises – swimming, gymnastics, fencing) [7]; c) the famous Mircea flag-ship becomes the first school-ship of a Romanian navy institution.

In its first years, the school was training only inferior ranks (non-commissioned officers), but once at European standards, the best graduates were directed to a military officers school at the special weapons division, or were sent abroad to the most important naval academies.

The 1887 regulation made clear the graduates were to be promoted to the corporal rank and sent in service of the fleet, and the ones that scored at least "15" were to be admitted without exam at the Military School for Artillery Officers, Engineer Corps and Sea Service (inside the special weapons division), stating that at the end the their studies they were once more going to be promoted to the lieutenant rank.

Between 1881 and 1891, the school had offered a small number of graduates: 40 (1888), 50 (1890), most of them being the commanders and the people second in command from the Romanian Fluvial Service which gave the navy personalities like: vice-admiral M. Gavrilescu, rear-admirals O. Nedelcu, N. Negru, Constantin-Niculescu Rizea, and commander N. Ionescu-Johnson. [8]

Starting the year 1893, through the decision no.745/4 of March, decision whose merits were mainly attributed to colonel Ion Murgescu, The School of Squadron Personnel opened its doors. The institution's two years plan had the goal of forming inferior rank people with diverse specializations for the fleet, people that couldn't come from the usual regiment schools. The number of students was set by the fleet commander and the formation plan involved: mathematics, geometry, Romanian history and geography, shooting stages, swimming, fencing, deck service and others. [9]

The education plan was commonly divided in two: a) the theoretical instruction (1^{st} of October - 1^{st} of April; 15^{th} of April - 1^{st} of May – exams); b) the practical instruction (1^{st} of June – 1^{st} of October, held aboard ships and with different programs for all specializations, ending with practical exams in September). After graduating the 1^{st} year, students possessed a corporal rank, and after finishing the 2^{nd} year they were promoted to sergeants. The 1^{st}, 2^{nd} or 3^{rd} grade certificate were obtained after sustaining certain exams announced in due time.

Due to the decision no.1093/26.02/1896 and as a result of establishing the Royal Fleet, a new educational navy institution was born in order to prepare the necessary personnel: a) *The Special School of Navy* (forming supervisors and professors for the different disciplines of military and commercial navy); b) *The Special School for Engineers and Experts* (training mechanics and masters); c) *The Application School for Navy Sub-lieutenants*.

The goal of this complex educational system was "to give special education to marine officers assigned to a specific weapon, upon graduating the military schools from all around the country". It was a two years stage, and its structure remained classic: a) the theoretical preparation stage (1^{st} of November – 28 of February: allocated for studying navigation, hydrography, combat maneuvers, torpedoes and electricity; with exams scheduled for March, from the 6^{th} and until the 15^{th}); b) the practical applications stage (aboard military navy ships: from 16^{th} of September to the 10^{th} of October). Students enrolled in year number two were commissioned to a rigorous exam for testing their skills in front of a jury that included the fleet commander, the marine arsenal director, a delegate of the war ministry, the school's commander, the ship commander and a number of professors as consultants. The minimum average score was "12" and no grade below "8" was allowed. The ones with a "14" or higher score were eligible for promotion ahead of the usual schedule and promotion's best students represented the top list for sending to foreign naval academies or fleets. [10]

Until 1909, the school held its classes in Galati city, next to the Fleet warehouse, and starting 1910 it was relocated to Constanta, next to the Sea Division. The number of graduates was relatively small. During 1903 and 1904 there were just three sub-lieutenants attending classes, and in 1905 – four.

Until the end of the XIX^{th} century and the beginning of the XX^{th} one, the personnel needed for marine army was completed with the help of other already established military schools: The Military Infantry and Cavalry School (with the special weapons division born in 1872), The Special School for Artillery and Engineer Corps (1881), the Military School for Artillery, Engineer Corps and Marines (1910). For teaching naval specific disciplines, the experienced officers were dispatched from the Fleet.

Sustained by press articles, parliament debates and public meetings, the need for a superior school comes to life in 1909, made possible due to the decision no.2928/29 of October, when the Navy Superior School opened it gates in Constanta, on Traian Street, location that today serves as the place of the Navy Museum. [11]

The educational process was adapted to our own navy needs. There were classes of military art, naval wars, combined tactics, ship fire and artillery, underwater defense, international navy rights and laws. During 1910-1911 there were five lieutenants attending classes, and in the next year, there were seven. The School's status was conferred by the Military Organization Law and by the decision no.4223/3 dating October 1912. In the sixth article it was stated that: "the navy school is a unit of direction and involves": a) The Superior Navy School; b) The Navy School for Military Experts; c) The Navy Firing School; d) The Hydrography and Piloting School; e) The Application Schools for Officers and Troops.

Although it had a short life - closed as a result of the First World War -, the navy education gave devoted and instructed officers to the fleet and country, people that proved their qualities and patriotic spirit during the military navy applications between 1916 and 1918.

4 The Navy School

After the First World War, Romanian Navy finds itself into the middle of a real crisis regarding personnel. If during the eve of this battle the number of pupils was reaching 189 (deck officers, engineers and mechanics), four years later the number was down to 141 due to the fact that between 1917 and 1919 the military schools didn't have any promotion of marine officers. [12]

Starting from navy's necessities and wanting to have institutions capable of preparing students at European standards, on 5th of May 1920 at the request of the Vth Division of Navy belonging to the War Ministry, the Navy section is separated from the structure of the Artillery, Engineer Corps and Naval School and Navy School is born.

This way, starting with the school year 1920-1921, through the ministry decision no.372/09.06.1920 and establishing decision no.674/7.07.1921, the system that trained navy personnel was reorganized into a new institution: The Navy School [13], having two sections: The Navy School and The Naval Institute. The Navy School was made of: a). The Training School (which represented a two years stage) – preparing the graduates having the sub-lieutenant rank; b) The Application School (with a one year stage) – perfecting the graduates for the lieutenant rank and then, putting them in service throughout naval divisions; c) The Master Class for marine captains. The Naval Institute had different components like: a) The Navy Children School and The Special Navy Masters School. Together they were the divisions of the Navy School.

Later on, by decision no.3434/01.10.1938, in order to have a better response to the evolution of the commercial fleet, there was established the "Maritime Trade Division", section which consisted of a three years stage for preparing deck officers, mechanics, radio communication specialists, agents and commissaries. [14]

The Navy School operated in Constanta, until the year 1948, on 53 Traian Street, in the former place of the Marine School. The School was also the first superior education institution to be found in the biggest city-port of our country. [15] According to the establishing decision no.674/7.07.1921, the goal of this Navy School was to "give basic naval education and instruction to the young people that wanted to embrace a career in the navy" [16], and the first internal regulation of the Navy School insisted on the "professional military, physical and moral preparation that was needed by the navy officers and the mechanic-officers of the war navy". The classes started on 13th of November 1920 with both stages (year one and two). The attendants of the second year (14) were selected from the graduates of the first one belonging to the naval division of the Artillery, Engineer Corps and Marine School. In the Ist year, after the admission from august 1920 there were 22 students enrolled. [17]

The ones that desired a place in this learning institution had to satisfy certain demands: admitted age was between 17 and 21, they were to have a not married status, they must had graduated a 7 classes mathematics-profile high school; also, another condition was to have Romanian nationality and no disciplinary sanctions in their backgrounds; candidates were expected to have a good health, to be presentable and coming from a middle to upper-class family. On their application, besides the obvious papers like birth and nationality certificates, study papers, the candidates had to present a written agreement of their fathers or legal tutors which stipulated that they will return the educational expanses if they would not serve the navy for 9 years. Agreement that is valid even nowadays.

Admission consisted of two written papers, a drawing test and an oral exam. The first written paper had two subjects from two disciplines at choice from different categories: plane or space Geometry, Trigonometry, Physics and Chemistry; the second paper was about History and Romanian language; the oral exam was again a confrontation with Mathematics, Physics and Chemistry.

For the school year 1924-1925, The Navy School had been structured on a three years plan: the Ist year – preparatory course, the IInd year - normal classes and the IIIrd year – application courses. [18] In 1928, The Preparatory School, separated from the Application School, has organized its education plan on two specializations: deck officers and mechanics. The Preparatory School for Mechanic-Officers emerged between 1929 and 1930 and had its first promotion in 1931. The graduates were ranked mechanic officers third class, the equivalent of sub lieutenant rank from the navigation section.

At the Preparation School, the education had two fields: theory and practice. The classes consisted of general military information and basic naval knowledge. These classes were organized in the following order: for the Ist year, there were exact sciences – Arithmetics, superior Algebra, plane and sphere Trigonometry and Physics - and for the IInd year they had descriptive and analytical Geometry, differential and integral Calculus, rational Mechanics, industrial Chemistry and Electricity. In addition, the Ist year was also studying Romanian language. Both years had special navy classes involving: costal and approximate Navigation, astronomical Navigation, maneuvers, marine machines, underwater mines and motorized torpedoes, naval artillery, naval regulations and maritime geography, naval application classes, techniques organizing gas services, naval constructions, Cryptography, Thermodynamics, Cinematics, underwater weapons, land weapon tactics. As common disciplines, we also have to mention the two foreign languages Italian and English and also Moral education.

The education plan for mechanic-officers was made of: Mathematics (Algebra, Geometry, Trigonometry, Analysis), Mechanics, maritime Geography and Oceanography, Navigation and Cosmography, maritime machinery, underwater weapons, naval constructions, ships theories, electricity, electrotehnics, radio communications, hydrography, weather forecast, ship maneuvers, topography, naval tools operation, boats, motor-boats, steersman and topman’s manuals, codes and signals, deck service, foreign languages, military regulation, Music, Gymnastics, Fencing and so on.

From the staff that worked at the Marine School, we should remember: commanders Ioan Georgescu (navigation, maneuvers, steering), Aurel Negulescu (navigation), Corneliu Bucholtzer (descriptive Geometry, submarines, mines and torpedoes), Petre Bărbuneanu (Geography, Mathematics, Meteorology, Hydrology, Oceanography), captain-commanders Victor Schmidt (ships maneuvers, submarines), Constantin Pogonat (Oceanography), captains N. Cristescu (navigation), Alex. Stoianovici (international law and navigation), lieutenant commanders Iacob Bălan (marine regulations), M. Constantinescu (Artillery, Firings and Ballistics), Alexandru Gheorghiu (Naval History and the History of naval wars). And the list could continue with former military and civilian professors which shined with their personalities in different domains: colonel (r) M. Ionescu Dobrogianu (Geography), he became corresponding member of the Romanian Academy, engineer Virgil Cotovu (differential and integral Calculus), promoted head manager of Constanta's port, engineer Gh. Rizescu, outstanding in the fields of mathematical analysis and superior algebra.

During 1921-1928 there were 158 attendants allowed in the Ist year and most of them, 140, had made it to the second one. Only 131 graduated with the sub lieutenant rank. Within the first promotion (1922) of the Artillery, Engineer Corps and Marine School, there were two personalities that must be evoked - Dumitru Ştiubei, remarkable marine painter and Eugen Săvulescu, which became the commander of the school. Even more, in the history of that school there is a memorable event that involved princess Ileana from the royal family. In 1928 she took the captain exam and was admitted with 8, being the first Romanian woman getting the certificate of deck officer first class. [19]

Practical classes used to begin on the 15th of May, after the theoretical education was over. This stage was held at sea, involving practical naval applications during June and July onboard Sea Division's school ships. They did exercises like: navigation, artillery, steering, signaling, managing sail and scull ships, training expeditions in order to recognize military and industrial objectives. After the Maritime Trade was born, practice was conducted also onboard their commercial ships.

Among Preparatory School's deck section's promotions' best, during interbelic times, we can mention some names like: Eugen Reţeanu (1922), Corneliu Lungu (1923), Ioan Bănică (1924), Haralambie Ringhinopol (1925), Gh. Stănescu (1926), Dumitru Mitescu (1927), Florin Bădulescu (1928), Ion Stoian ((1929), Constantin Costăchescu (1930), Tudor Brătulescu (1931), Nicolae Stoenescu (1932), Ioan Zaharia (1933), Mircea Constantinescu (1935), Marian Dumitrescu (1936), Ilie Stoian (1937), Gheorghe Popescu (1938), Constantin Panco (1939), Constantin Bucur (1940).

At the Application School, trying to develop the technical and tactical skills of the attendant officers, the education plan covered general military knowledge (aviation, links and communications, army tactics), naval classes (astronomical navigation, naval artillery firings, turbines, engines, mines and underwater torpedoes, applied electricity, submarines, naval theory, TFS, maneuvers, international maritime law), general classes (Mechanics, differential Calculus) and foreign languages (English and French). The graduation exam consisted only in subjects related to military navy: artillery, navigation, ships theory, under water mines, without any practical exam. This school gave Romanian Navy great names such as: Aurel Pelimon (1922), future commander of the Navy School, Gherghel Arpad (1922, the second series), Navy School commander (1943), Eugeniu Săvulescu (1923), future school commander (1944), Constantin Costăchescu (1930), the future commander of the "Dolphin" submarine, Grigore Martiş (1938), future commander of Military Navy, Nicolae Milu (1939), future commander of the Superior Military Marine School (1955-1958). [20]

Due to War Ministry's decision no.434/16.05 and their order no.81/13.04.1921, The Master Classes for all weapons fighting captains were becoming reality, having a five months stage and preparing the commanding staff and the Major State officials. The classes had vast topics like engines, submarines, TFS, navigation maneuvers, naval strategy and tactics, war games on maps, major state, battalion organizing and instructing, and history of navigation. The classes started with 7 captains, number that doubled next year. Among the professors, there were famous ones like: Ioan Izbăşescu, Eugen Roşca, Gheroghe Koslinski, Corneliu Bucholtzer and so on. The 1922-1923 promotion also leaves us names like: Iacob Bălan, August Roman and Horia Măcelariu.

When classes were over, the attendants were sent on navy ships (destroyers, a.s.o). Graduation papers were read by demanding juries and followed by artillery firings, maneuvers trials and navigation applications at sea. Only 132 from 176 marine officers had finished the different stages of preparation.

After war emerged, the Navy School had to relocate and adapt itself in order to suit the new situation. It functioned in Turnu Severin, Tariverde, Zimnicea and Turnu Măgurele. It was able to accomplish its goal only by the devotion of its commanders: Alexandru Bardescu (1939-1940), Vasile Petre (1941-1943), Alexandru Stoianovici, Gherghel Arpad, Aurel Pelimon (1943), Eugen Săvulescu and Ion Borcea (1944).

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THE ACTUALITY OF GUSTI'S ANALYSIS REGARDING DICTATORSHIP

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Abstract: *Familiar with the interbelic Germany's political life, the famous professor was preoccupied with finding explanations and arguments, both of scientifical and common sense nature, regarding the unfortunate ascension of Nazism and its leader in the history of Europe and humanity. In a certain way, the explanation for the appearance of this phenomenon or of the totalitarian leaders (dictators) resides even today in the contemporaneous political theory. It becomes the starting point in analyzing political speeches which have persuasive and manipulatory valences, and also in the depiction of a taxonomy for political leaders.*

Keywords: *dictatorsschip, Gusti, culture, leader*

1 Gusti's personality and the German culture

The scientifical and pedagogical personality of D. Gusti has been formed under the undoubtful influences of last century's German culture and education. In his initiatic journey, marked with study stages at great universities - Berlin University, Leipzig and Halle ones, and also The Criminalistic Seminar from Charlotenburg -, D.Gusti had the privilege of having contacts with the great European spirits of that time. He followed classes held by: F.Paulsen (philosophy), G.Simmel, W.Sombart (sociology), A.Wagner, G.Schmoler (economy), Dilthey (historiography), Bastian (ethnology), Wund (psychology), Ratzel (geopolitics), Bucher (political economy), F.Tonnies (sociography), K.Lamprecht (history of culture), Franz von Liszt (criminalistic sociology), Stammler (social philosophy) and the list could continue. The final test in the formation of his multidimensional personality was to obtain his ph.d in 1904 at the University of Leipzig, having the thesis of "Egoism and Altruism" and sustaining it in front of a jury formed by W.Wund (president of the jury and coordinator), K.Lamprecht and K.Bucher (members of the jury and examiners).

After almost three decades, the former student - now having the professor status - is invited to give a speech in front of the academic world and the universitarian youth of Munchen, Berlin and Leipzig. Gusti taps into the new cultural, social and political world of Germany belonging to the '30s: a clouded world that became ambiguous, sad and dominated by the ascension of national-socialism and its infamous leader.

This journey through Bavaria (Munchen), Prussia (Berlin) and Saxonia (Leipzig), gave him the opportunity to meditate on and study the Hitler phenomenon [1], to formulate a viable answer to a common sense question: "how can we explain that a man - not being even a

German citizen, with no past or culture -, that started as a simple corporal without connections and political party, can dominate the people and get such an ambition of ruling the whole world." [2]

In order to find accurate answers, the famous sociologist goes deeper into the widely-spread theories of several thinkers of that time - W.Schopenhauer, Ch.Renouvier, Lombroso, Brentano, E.Eaystrow, Hoch Lange-Eichbaum, Kretschmer, Plaut, Oswald - thinkers that studied the human personality in every direction, from character to talent, to the state of accomplishment or that of genius. This search, of a scientific interest, was meant to confirm or not his own theory about the social personality. For Gusti "the famous problems concerning the philosophy of history and the great man sociology - both have found in Hitler a clear case - one of the most eloquent -, showing how to create political cultures. Through this case, the controversial great man theory has been validated".

2 Gusti's analysis regarding the birth of the Nazi dictatorship

Hitler's role, Gusti used to mention, "cannot be scientifically determined by itself, but only when we see it related with the world he lived in".

The ascension of the Nazi political leader must be regarded in a certain economical, social and political context which marked that time in history: an inefficient republic lead by a mediocre social-democrat cabinet that came with the German revolution. The socialists' promises - the socialization of production means, a society without layers, good care and wealth for average and poor people - were still just words. Economically speaking, Germany was just one step to collapse, situation due to the high unemployment rate (7 millions), the financial bankruptcy, the depreciation of national currency and the growing inflation (one dollar was exchanged for one thousand billion marks), not to mention the general poverty. From a political point of view, the "new system was living only through arrangements" and excelled in inefficient birocracy. The govern was formed by mediocre persons which lacked political will and culture; there were just demagogic parties; the army was exhausted after its defeat in the first war; the society was in a moral crisis and psychical depression with the youth having an unclear and unsafe future. The German republic was born in tragedy, in 1918 and ended in 1932, also in tragedy. With the 1933 vote - Hitler (11milions) and Hindenburg (10milions) - the German people showed the desire for a solution to satisfy their needs and future ambitions and also to get out of their difficult situation.

The Nazi saga is depicted as a drama in five acts: act I, the modest man, without party and suffering; act II, his political birth; act III, the fight and strategy to conquer power; act IV, the fight and strategy to conquer the world; and finally act V, the collapse and his suicide. [3]

Hitler's success – obtaining power and becoming the most powerful man – can be explained not by his own qualities, his fascinating and seductive speeches, but in the first place should be regarded in terms of economical, social and spiritual conditions which determined this social psychological phenomenon. Gusti concluded "the great man, Hitler, was nothing else but a general exponent of his time, a picture of all the errors, contradictions and aspirations, an exponent of the supranationalist theory – the racism – which carried the horrifying label of unexplainable crimes and trinity – domination, extermination, servitude". [4]

Gusti's sociological analysis closely followed the next principles: a) the usage of all empirical means of critical and syntethical knowledge regarding the surrounding reality; b) putting science in social action's service; c) putting social action to help the personality creation mechanism; d) making the personality to create values; e) prioritizing values and integrating personalities in vaster ones, through unification of their goals and coordination of their means; f) the substitution of individual effort through organized cooperation of personalities in institutions; g) putting elites in masses' service. [5]

This case-study, offered the Romanian sociologist the opportunity to display a real methodology regarding social psychology, treating the ascension to the status of great Nazi man according to a recipe that even if it's no longer viable, through extrapolation it still can be applied in the process of masses manipulation and domination by their political leaders or the now-existing power structures, no matter their color. The infamous Nazi leader sustained that masses are to be dominated, amused and that was what he did.

Following this idea, the author writes about the techniques and methods used by the Nazi leader in order to manipulate the German people: a) dentification with the will of the people; b) the spiritual uniformization of the whole nation by the national-socialist party (through constraints, persuasion, terror, educational propaganda made possible by mass-media, specific institutions and the transformation of universities in tools serving spiritual domination); c) he party identification with the state and the nation; morphing the national-socialist party into a state-party (the exercise of power upon all activity fields, the state-party being all over and all powerful); d) he youth continuous political education; young people were raised with one goal, that of executing Fuhrer's orders; education, following these rules, was supposed to target physical strength and shape characters towards obedience. There were established certain "comrades fortresses" and only their graduates were admitted to teach at universities, selection based not on the academic merits, but on people's political loyalty. Universities were simple means of executing orders and measures sent by the Instruction Ministry; e) ropaganda actions were based on foolproof psychological and sociological recipes; the gatherings were impressive and treated as real sociological laboratories; they were representations of sounds and lights, with theatrical character and elements of optical illusions. The violence of speeches, the inducted contagious fanatism, the hysterical uncensored feelings and obsessive themes like Jews, Germans, peace, socialism, community and nation, all these made sure the meetings' success).

Trying to explain the efficiency of Nazi speeches following the Goebbels sequences, Jouvenel finds five components: the spotlight (the teller identifies an existent situation and turns the attention towards it), the judgement (the sentence which is not favourable to the situation and calls for action), the perspectives (the speaker depicts a future better then the present, future that becomes target; the better future is depicted as opposed to the worse one; the two possible future scenarios which are one against the other induce hope and fear), the steps (shows the conditions and the means for fulfilling the better future) and the imperative must (the better future which was just an alternative becomes mandatory) [6]; this type of speech combines in a perfect composition the moral aspects with the pragmatic ones, given that the latter part motivates and leads to action; the governance methods (torture, assassinations with the "kill! to be rewarded" motto, all the means are premises to completing the task; justice being replaced by politics, and the law with the Führer's will).

Although different through their manifestations influenced by place, men and people, dictatorships have some common background regarding their origin and behavior: a) tey appear as salvation actions (after restless and social uncertainty years, in order to stop chaos and create a better life; even if some dictators have been good managers and have favorized arts and literature, all of them have eventually left behind only disasters and pain); b) this kind of regimes lack juridical support, having decisions unknown to the public and no transparency at all; c) they start as temporary administrations but end up being permanent; d) dictatorships are characterized by intolerance; e) to self sustain, such regimes use certain ideologies of justified religious inspiration having specific rituals: the cult for admiration and idolatry, the unconditioned obedience towards the one leader; f) dictators always need two companions: the nation (which they dedicated all the glory, power and reforms) and the state (being regarded as god, as dictators are seen as students of Machiavelli and Hobbbes); g) dictators have unlimited freedom, having to answer only to themselves. [7]

According to Stephane Courtois from the *Communisme* magasine - also the person responsible for the *Democratie ou totalitarisme* work -, the totalitarist regime is an extremely logical political system, even if it belongs to a logical frenzyness. Communism, fascism and nazism are logical: they are all based on an utopic and scientific vision of the world and on the revolutionary action theorized by Lenin in "What's to be done?". Hitler brings rasism as a scientific truth, thus becoming the major element of his politics. Beyond its human monstrozity, the industry of the final solution underlines the totalitarian logic.

3 From the "great man" theory, to the one of the totalitarian leader

The genesis of Nazi regime is completed with the great man theory, a pertinent x-ray of Hitler's personality and his cult.

For this reason, Gusti brings more explanatory arguments by introducing a new variable: the psychology of people. The character of German people, in his opinion, was made of two dimensions: the soul (belonging to the highest spheres of the spirit: the world of transcendental philosophical ideas, the philosophy of history, universal science and music) and the automatic (that loves hierarchy, ranks, blind discipline, military, uniforms and obedience towards authorities). This idea is formulated again, nowadays by Almond and Verba, regarding the behavior of different people when it comes to freedom values, reaction also commented by S. Huntington: Germans display an affinity for rules and respecting the rules, discipline being regarded as value that outstands freedom. [8] A possible explanation for the paradoxical evolution of German people, from the remarkable contributions in the development of human civilization and culture to the atrocities of Nazism.

Hitler was a controversial personality; it was a man lacking intelligence or culture, belonging to the classic fanatic and autocratic type (one step higher than even Wilhelm the

Second), abnormal and psychopath (this being the opinion of his own major state chief – Haler), but also a great actor, experimenting social sociology and psychology, caring a simple speech with undeniable power of suggestion on the crowd, mastering the propaganda mechanism. Gusti observes that *Mein Kampf*, the national-socialist bible, was made of common things, of subjective judgments, superficial lectures and filled with many quotes from the old dr.Lueger. Hitler's power of suggestion and persuasion was fueled by his status of Germany's son and exponent. A poor people, in despair and without self consciousness, would follow the one that represents a more prosper society and a dominant place in the world for their country.

His ascension is tied to the magic of the meetings. In order to impress and dominate the masses, there were theatrical shows with directors and mise-en-scène, shows of applied arts. The meetings were held after a very precise algorithm: an architectural part, a theatrical and a musical one, with drums entries and exits. This way the meetings had the looks of grand celebrations and the electoral campaign was transformed into a "Wagner opera, with flags, orchestra, uniforms, songs, hymns, frenziness (…)". This picture was completed by a certain type of speech and beyond the well made speech there was the man standing in front of the crowd, with his voice inflexions. The speeches followed the same rule: a slower start, with pauses and modulations on a sarcastic tone in order to make the people curious. Gradually the sentences got louder and powerful and the speech ended in violent and vindictive tones. "He was the leader that knew to conquer the crowds, to rule the people through the magic of a psychopath verb". Eyewitnesses remember that when Hitler was speaking, the walls needed to be solid and his eyes were turning on and off like the headlights of an automobile. Hitler's cult was based on the naivity, fear and despair of German people being in the middle of a crisis.

Today there are a lot of documents that refer to this subject. This interest is determined by a reality that the author observed while being preocupied with the study of the power theory: "the dictatorship problem is not something new, but it belongs to all our history, as an illustration for the central problem of sociology and ethics, of the ballance between individual, society, nation and state on one side, and the authority and freedom on the other". [9] And as long as the totalitarist systems still represent a threat, the study of this matter is justified.

The totalitarist leader is illutrated by the "one leader", "hero leader", "parent leader"; in all its shapes it is taken for the one that is the most influent and whose wisdom and energy bring the masses an unlimited trust in him. Mussolini, Hitler, De Gaulle, Lenin, Stalin, Mao or Ceauşescu have one thing in common: a charismatic authority. Even there are points of view that deny this approach (including Lewin which puts the accent more on the terror mechanisms), we can't deny the birth circumstances of these personalities: crysis situations of economical and political nature, the lack of social perspective. They have build and justified their ascension by using the mystical hero image, being the saviour, the men that erase peoples' uncertaintes, giving them hopes to a better future following the wellknown golden age model.

Hitler represents the person that cannot be refused, the visionary, symbolizes the good will, the historical chance to a better society. The crowd gives absolute power to the dictator, having arguable criteria: the years spent in prison or exile in the precedent regime, the rezistence in front of the enemy and so on. Communication with the crowd is episodic and has a mystical element. His public appearances have a worship athmosphere. His speeches represent true means of knowledge (the red book of Mao, the green book of Geddafi, Hitler's Mein Kampf) and become law for his followers. The party propaganda, through different mottos – "Believe, obey, fight", "Mussolini is always right!", "The party is always right!", "The most loved son", "The Party-Ceausescu-Romania!" – has amplified people's obedience in front of the leader which has an autoritarist personality, also being a very seductive warrior.

The totalitarist leader - characterized by Arendt, Hayek, Revel, Dahrendorf, Pasquino, Linz, Betea, Tismăneanu and others – is the "personification of the ultimate sense", "the supreme leader", the one that cannot fail, that establishes goals on this own, the one that transforms his followers into fanatics. An unpopular leader is a leader without power, but the power of a totalitarist leader is unlimited. The unconditioned adherence of the party to his person and projects is gained through a variety of means: the prophetic speech, the will for power, the claimed dependency between masses and leader, the total control, the demagogy, the arrogance, the populism, the strong belief that he represents the majority, the utopic ideologies and sometimes even terror (psychological and political). [10]. H.Arendt notes that "The will of the Führer represents the law for his political party" and the whole organizational structure has a single purpose: spreading the leaders will to all party members.

"The first goal of a leader in a totalitarian regime, is not that of governing through force or that of destroying his enemies, but of making its followers to honestly think like their leaders" (B.Ficeac). [11] Hayek completes the picture by saying that this kind of leaders can be distinguished by the lack of ideals, beliefs, or moral principles and through their negative platform. In order to maintain the mystic of their own merits ascension, the biographies of communists, nazists or fascists leaders insist on their humble origins (Ceausescu – son of peasants with no occupation; Mao Zedong – his father was a teacher; Lenin – his father was a school inspector, after finishing law; Hitler – had a custom officer father and an unemployed mother; Stalin – his father was a shoemaker and his mother was without a job; Mussolini – had an iron worker father and a teacher mother) and also on "the disaster of their personal lives". [12]

The communist leader, species belonging to the totalitarist and autoritarist leaders, represents a foolproof example, well exercised throughout the history: the forgery of his biography (he is given credit for initiating and elaborating the strategy regarding the most great deeds in the recent history of his country), the leader's association with the party, people and country, the projection of his saviour image (the one that brings prosperity to the masses, the follower of national historical traditions, and also a great army commander), him taking care of people's destiny - a man of continental and world political value. [13]

To conclude, the up to date totalitarist leader model is not too different or not at all then the model used by nazism a few decades ago.

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SOME ELEMENTS CONCERNING THE ASSESSMENT OF THE TRAINING AND THE CAPABILITIES TO PERFORMANCE BY PHYSIOLOGICAL TESTING OF SWIMMERS

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Abstract: *This paper is presenting some assessment of the specific training level by physiological tests on swimmers. The work is inspired by testing protocols outlined by Australian Swimming Inc. Many of the protocols have already been described in specialized works in order to suggest various fitness compounds for highly trained swimmers. In the same time, even if some national level swimmers would be asked to stand some special tests it doesn't mean that would stop the creativity and inspiration of coaches and scientists to use their own methods and testing systems. We are only presenting here general issues stressing on scientific methodology in order to help the practice action.*[1]

Keywords: *swimming, physiological test, submaximal aerobic/anaerobic effort, assessment, training level, performance*

1 Introduction

Swimming is a sport that presents significant opportunities to apply information gained from physiological testing.[2] There are a number of benefits that can be obtained from a well-planned testing program.[3] Within Australia, the testing protocols outlined in this section have been The national body (ASI) is administering these tests to Australian swimmers through its National Team and National Event Camp Program and in High Performance Center linked to state institutes or academies of sport.[4] While some national level swimmers have a formal requirement to undertake these tests, it is not the intention to stifle the creativity of scientists and coaches in developing and utilizing their own swimming testing protocols.

The tests outlined in this section are largely pool-based for reasons of specificity and practicality. Few nations possess a purpose-built swimming flume necessary for routine measurement of oxygen uptake during swimming. Although it is possible to measure oxygen uptake during swimming in a conventional pool, or to estimate it using techniques such as backward extrapolation, the practical difficulties associated with these approaches have limited their use in testing large numbers of swimmers.

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[1] Col. (Rs.) Dan Nicolau also contributed to the achievement of this work.
[2] Treffene 1979, Troup 1984 and 1986, Prince 1988.
[3] Refer to chapter 17 Protocols for the Physiological Assessment of High Performance Track, Road and Mountain Cyclists.
[4] Approved by the Sport Sciences Advisory Group of Australian Swimming Inc (ASI).

A large number of different testing protocols have been described in scientific and coaching literature to assess various components of fitness in highly trained swimmers (Roberts 1991). While these tests only provide an indirect estimation of the contribution of different energy systems, they do not have the limitations inherent in laboratory tests of running, cycling and rowing where results must be transferred to the field. The underlying rationale for the tests described here is this: physiological information derived from a swimmer during sub maximal and maximal testing provides critical detail for the coach to determine training loads and monitor performance improvements (Roberts 1991).

2 Administration of physiologic tests

It is recommended that physiological testing of national and international level swimmers be conducted approximately three times per competitive cycle, based on the preparation for the major championship in each calendar year. In normal circumstances there are two competitive cycles each year. Most leading swimming nations conduct their national championships in March and April each year, with the major international swimming championships held in July and August (summer in the northern hemisphere). Testing dates should be organized at least one full competitive cycle in advance.

Based on an average 12-16 week preparation testing should be conducted at the following points:

- early in the preparation (2 weeks into preparation: 10-14 weeks from competition);
- mid-preparation (6-8 weeks from competition);
- pre-taper (3-4 weeks from competition).

It is suggested that, where possible, the tests should be conducted during a recovery or adaptation week to avoid the interference of residual fatigue. Testing should be standardized so that the same day of the week (e.g. Monday, Tuesday) is used each time. The following example is suggested to illustrate a typical testing schedule (Table 2.1.).[5]

Table 2.1

Day	Saturday	Sunday	Monday	Tuesday
AM	Normal Training	Off	1. Blood Testing; 2. Anthropometric test, then, Normal Training	Normal Training
PM	Normal Training	Off	3. Aerobic Test; 4. Speed Test.	5. Stroke Dynamics; 6. Speed-Endurance Test.[6]

[5] **Acknowledgement:** David Pyne, Wayne Goldsmith and Graeme Maw - *Physiological Test for Elite Athletes* (Australian Sports Commission), Editor: Dr Chris Gore, Publisher: Human Kinetics, PO Box 5076, Champaign Illinois, 2000; Swimming Chapter is 27, pp. 372-382.

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3 A short introduction to australian swimming

Swimming is a worldwide and popular sport, where you can participate at any level. Some enjoy the sport for fitness, some for recreation purposes while others compete.

Competitive swimmers may cover 10,000 to 14,000 meter a day, 6 to 7 days a week. Becoming a successful swimmer takes time, skills, hard training and a love for the wet element. Different techniques are used - front crawl, butterfly, backstroke and breaststroke - and preferable distance varies. Swimming events rage from 50 m (takes 22-26 sec.) and 1500m (takes 15-17 min.) Open-water or long-distance may range between 1 km (10-12 min.) to 25 km (5-6 hrs.) (Australian Swimming Inc. 1996) Shoulder problems and injuries are common in swimming, due to high repetition rate, extreme range of motion and the force required for propulsion. However, swimming is a very aerobic activity and enables people with musculoskeletal problems to train, and avoid impact forces like in closed chain activities (Zachazewski 1996). The focus of this paper will first and foremost be on the physiological testing of the swimmer. A shorter overview of the sport and its demands will be given as well.

4 Some factors influencing swimming performance

Sex Differences: Males tend to swim faster than females. Women have higher percentage of body fat than men, whereas men have more muscle weight. These results in women floating better and showing a greater swimming economy, 30% lower energy cost than men have been reported (McArdle, Katch and Katch 1996).

Strength Swimming power and especially upper body strength have been demonstrated to be crucial to success in sprint swimming. 86% of one's performance in a 25m front crawl sprint result from the swimmers' strength and the ability to develop power. For the competitive distance swimmer the strength component is less. At 100, 200, and 400m, the contribution of muscular strength drops to 74, 72 and 58%, respectively. During slow, low-intensity swimming most of the muscle force is generated by slow twitch fibers. As the muscle tension requirements increase, the fast twitch fibers are incorporated. In sprint events (50-200m) demanding maximal strength, the second group of fast twitch fibers sets in. The tendency is that swimmers have higher percentage of slow twitch muscle fibers in their shoulders and particularly musculus deltoideus. However, muscle fiber composition appears not to be a deciding factor in successful competition. Swimming is performed almost totally with concentric contractions (Costill, Maglischo and Richardson 1992).

5 Dynamic strength and swimming performance

Dynamic strength is an important determinant of swimming performance. Studies have found that upper-body muscular strength and/or power output correlate highly with swim velocity over distances ranging from 23 to 400m. Also, swim and swim-specific resistance training (e.g. bio-kinetic swim bench training, reverse current hydrochannel swimming and in-water devices that the athlete push off from while swimming) improves a competitive swimmer's velocity in events up to 200m.

This training can result in improved stroke mechanics, such as stroke force and distance per stroke. Research tends to conclude that stroke mechanics may be more important in determining velocity and swim success than upper body strength (Tanaka and Swensen 1998).

6 Energy generation for performance effort

The amount of energy required to swim is related to the intensity and strokes used. The demand for energy is reduced proportionally with the skill of the swimmer. Traditionally, monitoring the oxygen consumption during sub-maximal swimming, at speeds below those in competition has been the measure for energy use. However, swimming performance seems to be more dependent on the skill of the athlete than the $VO_{2\ max}$. values (Costill, Maglischo and Richardson 1992). Swimming fast is a matter of increasing propulsion while reducing the resistance of water to forward movement.

In swimming, energy is consumed both to maintain buoyancy, to generate horizontal movement through the use of arms and legs and to overcome drag forces in the water. The speed, size and shape of the swimmer and the fluid medium result in the degree of drag. The total drag force that the swimmer must encounter consists of wave, skin friction and viscous pressure drag. In comparison to running, swimmers have to use four times as much energy covering the same distance. In order to reduce the body drag, optimal technique especially of the arms is essential, and the use of wet suits have been introduced and show to decrease the body drag by 14%. *"Elite swimmers can swim a particular stroke at a given velocity with lower oxygen uptake than relatively untrained or recreational swimmers."* That is, skilled swimmers use more of the energy produced per stroke to overcome drag forces. And secondly, they cover a greater distance per stroke than untrained, who use energy to move water (McArdle, Katch and Katch 1996).

7 Specific endurance

Endurance is defined as the ability to repeat muscular contractions without fatigue. Fatigue is a decline in the level of performance. The performance relates to muscular contraction and depends upon different energy sources. This again must be reflected in the duration and intensity of the contractions, and tests must seek the energy source used for these contractions (McLatchie 1993).

8 Aerobic endurance

The ability of the body to support oxygen to the working muscles and to extract waste products which are transported with the bloodstream, to the airways is defined as aerobic endurance. Traditionally the Maximal Oxygen Uptake Test ($VO_{2\ max}$) described by Sinning in 1975, has been applied to measure aerobic capacity (McLatchie 1993).

9 Lactate endurance

This is the improving capacity, the ability of the muscle to rely on anaerobic metabolism at any given points of sub-maximal intensity of exercise. Lactic acid is the end product, and muscle fatigue will occur if it is not removed (McLatchie 1993). A study by Kesinen et al (1989) looked at different modes of lactate tests. They investigated swimming velocity, and blood lactate and heart rate responses with varying durations in separate swimming loads, and found that the most accurately evaluation of anaerobic capacity was using 2 × 100m or n × 100m modes (Keskinen, Komi and Rusko 1989).

10 Anaerobic endurance

From another point of view anaerobic endurance is the ability to maintain high levels of work intensity (McLatchie 1993). The Anaerobic Threshold is the highest work intensity beyond which lactate begins to accumulate in the blood. The metabolic acidosis occuring above anaerobic threshold contribute to limitation in performance (Cellini, Vitiello, Ziglio, Martinelli, Ballarin and Concorni 1986).

The aerobic training zone is differentiated into low-intensity and moderate-intensity efforts. This zone extends until the rate of lactate production exceeds the rate of removal, where the rate of lactate accumulation rises over the baseline, is termed the aerobic threshold. Once the threshold is crossed, the swimmer will not be in steady-state and cannot continue swimming at this pace for an extended period of time. At this stage hypo-aerobic fibers will dominate compared to hyper-aerobic fibers. Training over the aerobic threshold will affect the muscles ability to tolerate or buffer acid, and to remove lactate from the intra-cellular environment (Australian Swimming Inc. 1996).

11 Energy system contribution

The intensity and duration of swimming determines the relative contribution of the anaerobic ATP-PC and lactate energy systems, and the aerobic energy system. "In the shortest swimming event, the 50m sprint, the relative contributions for each of the systems are: ATP-PC 65%, anaerobic glycolysis (lactate) 30% and aerobic 5%. For a 200m event the contributions are; ATP-PC 10%, anaerobic glycolysis 20% and aerobic 75-80%. Open water or long-distance events almost exclusively on the aerobic energy system" (Australian Swimming Inc. 1996).

12 Physiological testing of swimmers. the contribution of the coach

The coach is an important person with ability to observe and make assessments of performance, but there is also a need to have objective measurements, being more valuable in giving some dimension to the result, e.g. time, distance, score (McLatchie 1993). The physiological tests are designed to follow the swimmer's physical capabilities, improvements achieved, and to assist in planning the training program (Costill, Maglischo and Richardson 1992). Total fitness consists of strength, speed, flexibility and endurance, and all of these aspects should be evaluated.

The following points should be considered before assessing fitness:

- pre-test procedures;
- the purpose and intention of the test;
- the suitability of the test and the equipment used;
- the statistical criteria for the test;
- the use of the test results (McLatchie 1993).

The normally constant environment the swimmer trains in and the high degree of control that the coach has over the volume and intensity of training workloads, makes swimming an excellent model for applying information gained from physiological assessments (Draper et al. 1991). Information gained from testing during sub-maximal and maximal modes provides critical details for the coach to determine training loads and monitor performance improvements (Australian Swimming Inc. 1996).

The characteristics of a highly-trained swimmer are of high power and endurance. Endurance is related to the power and capacity of the aerobic energy system. This can be assessed indirectly with a graded incremental swimming test (7 × 200m step test). Power is a product of strength and speed. The use of a maximal effort 25m performance test (2 × 25m speed test) is applicable for power measurement. Muscular power may be defined as the power and capacity of the two anaerobic energy systems (ATP-PC and lactate energy systems). Speed and endurance can be estimated together in a 6 × 50m test.

Swimming is a very technical sport, and assessing the technique or stroke mechanics is important. A 7 × 50m incremental test, where stroke mechanics can be assessed from sub-maximal to maximal speeds has been suggested.

These four physiological tests have been suggested by the Australian Swimming Inc. to be performed by every swimmer. There have been presented here only several modes of testing a swimmers' fitness. Other tests that can be applied are blood testing, anthropometry, heart rate measurements, start test, turn speed test, strength tests, vertical jump test and subjective rating of effort. Three physiological tests of swimmers - one aerobic, one anaerobic and one ATP-PC test, will be described in detail later. A short description of stroke mechanics measurements will also be included (Australian Swimming).

13 Conclusions

Even if some national level swimmers would be asked to stand some special tests it doesn’t mean that would stop the creativity and inspiration of coaches and scientists to use their own methods and testing systems. We are only presenting here general issues stressing on scientific methodology in order to help the practice action.

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THE NEW ARCHITECTURE OF ECONOMIES' TYPOLOGY INTO GLOBALIZATION CONTEXT

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Abstract: *Over viewing the most recently evolutions throughout global economy, we can easily conceive that the collateral effects of economical globalization and market integration, represents the main issues debated in specialized professional or political circles. The first step toward regain the global markets functionality is to review as a sine-qua-non condition, the institutional and functional structure of financial system and global economy system as well. In such context, this paperwork is meant to propose a new architecture of economies' typology, reflecting in fact the most recently particularities of markets' function. The criteria took under consideration has been the relevancy related to commercial and financial flows. Even the parameters presented here are quite abstracts in lack of detailed statistical data are important to reflect the new causality tied between economies in the new context of globalization.*

Keywords: *globalization, global economy, economical integration, global factors*

For most of the specialists preoccupied by the globalization phenomenon, the concept of the economical globalization induces, first of all, complexity due to the subjacent political or social-cultural reasons, the globalization being most of the time associated, on a perceptive level, to its implications. Starting with the nations relevance to the static and dynamic panel of the economical globalization, in addition to the classification recommended by the OCDE, there can be defined as relevant towards this phenomenon effects, three categories of economies, such as: economies based on an offer (offer side economics countries), economies focused on the internal demand and the export of capital (supply side economics) and finally the economies based on the energetic resources supply. This classification is based on the weight and the relevance of the nations within the international flow of goods, capital or labor, the characteristic features correlative between these economies, being used to define the dynamic interdependencies settled at the global level (figure no.1). First, within the category of *based on export economies* are included all those countries that built their economical priorities on the account of attracting capital investments as to capitalize the competitive differential resulted from accepting some social costs (income differences, standard) or environmental costs. In fact this category is adopting as a long term strategy the priorities of supply side economics theory.

Leaders of this echelon are the developing countries (emerging countries from South America and from South – East Asian area), followed by the undeveloped countries (possibly considered ‘witness’ countries). Their rhythm of economical growth is due to the production of goods with a lower rate of processing labor costs and a lower added value as well. The competition is actually guaranteed by the reduced cost of the products or services acquired/obtained by these countries which either accept to preserve social conditions under average (low salaries/incomes, reduced social assurances, improper conditions of health and work), deliberately assuming the effects of environmental damage as a price of relocation of the polluting production coming from the developed countries. These countries have an important weight in the world/worldwide trade and a constant rhythm of economical growth, based mainly on the export of raw material and products with a lower or medium rate of processing, based on the production in lohn (indirect export of labor) or franchise.

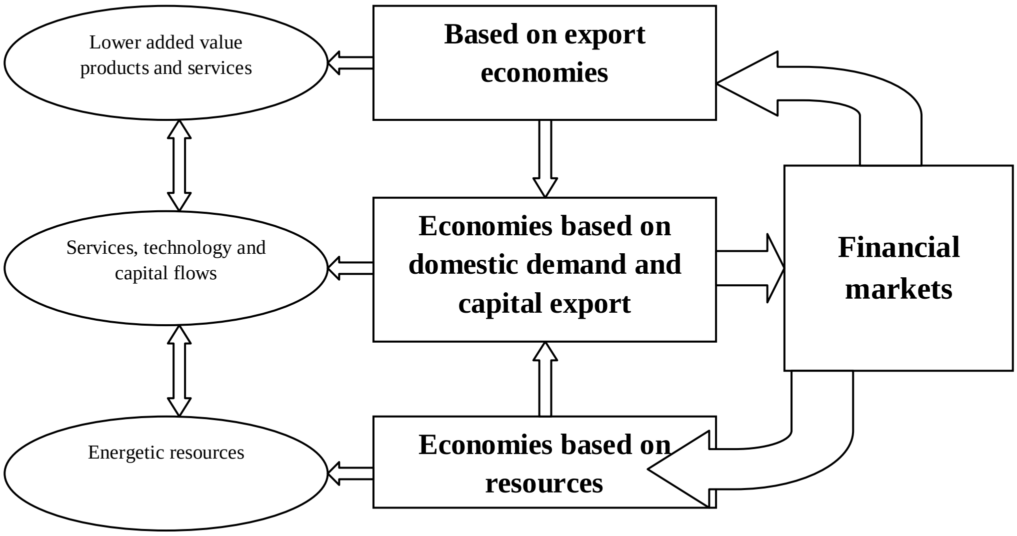

Figure 1

The economies based on offer, count on their economical growth mainly on the export to the countries with high incomes per/capita neglecting the domestic consumption. Thus, they draw the main benefits, accumulating higher incomes from external market. But the volatility of this kind of economies is major, comparing with the volume and the structure of the necessary capital to finance business. Also, the evolution of the balance sheet of payments of these countries is directly correlated to the level of demand formulated by the developed countries. On the other hand the accumulation is focused as a waste to the national reserve, preserving in this way the capacity for borrowing the international liquidities. But the national reserve, nominated in strong currency helps practically the high deficits oh developed countries.

The *economies based on domestic demand* category do not necessarily belong to the countries known as consumers above all but they mainly represent that specific class of countries, strong developed focused importing goods for an exclusively cheap consuming from offer based economies. Therefore, these economies draw benefits from a superior power purchasing parity and from the ability to speculate the price differential in terms of labor costs. The manufacture of a base product within a country where the guaranteed average income per capita outruns to ten times the average income per capita on the global level, obviously leads to a lack of competition, taxed by the unequal rules of the market economy. The production relocation tendencies, triggered by lower added values, from the developed countries to the developing countries, enable the growth of the demand from developed countries to a constant, diminishing the price and maintaining the inflation within reasonable rates. In this way, it is being assured the comfort of the daily consuming chimney and the increase of the life standards speculating the investment deficit from the developing countries. On the other hand, the countries based on the internal consumption have specialized in exporting goods or services with high added value, exporting the capital through the financially markets or through the banking system.

Services based on the informational and communicational technology and also on the financial and insurances industry have no comparison when we speak about developed countries. The export of licenses, patents, engineering or know-how procedures practically are monopolized by developed countries, the advantages being selectively ascribed to the capital suppliers. In spite of their own domestic demand increase, based on imported goods, the incomes of the developed countries become major throughout the export of services perspective, maintaining a monopoly thanks to the substantial added values.

The deficit of the balance sheet of charge account can actually be found within the foreigners assets, incorporated to sustaining the current account, sustaining or attracting investments. Losses within the current account imbalances resulted can be reduced by the developed countries either by moving the incomes obtained from the allocated investments efficaciousness or by reducing the internal demand or reorienting it towards its own production sources. The technological monopoly practiced by the developed countries with a 'white' and healthy economy, has a double valence: on one hand it offers exclusivity over the benefits of the technological progress trade (known as the most profitable neo-factor of production) and on the other hand, it assures a strategic priority within the process of globalization.

Due to the new international climate, taking advantages from the investments accumulated during the last decades, the developing countries, despite the deficits of their own capital accounts, increased the quota in international flows export, becoming the most important suppliers of investment resources for the developed countries. Thus, the developed countries maintained their supremacy in defining the structures of the direct investments on the international level, meaning the increase in capital export of the other categories of economies Yet, as a research groups of UNCTAD estimated, the tendency on a medium and long term will be defined by the returning of the capitals leaked out toward the developing countries, back to the mother countries, monopolizing the credit segment.

The trend of modifying the balance sheet of capital in developing countries favor, that have the immediate chance of becoming exporters of capital, does not necessarily represents a positive effect of the economical globalization and is not compulsory bring a major benefit to this kind of economy. At least as long the technological monopoly belongs further to the developed countries. The 'leak' of the capital out of the developing economies, despite the UNCTAD specialists positive opinion, proves the incapacity of the least developed countries to redraw their work profits within the reinvesting procedure on a domestic level. Moreover, the major deficits recorded by some developed countries in terms of capital account (e.g. USA, more than 600 billions USD), certifies the major contribution brought by these economies in the financial integration of real economy. These countries are actually giving away the surplus produced on the account of the technology export and on the outturns cashed as a reward for invested capital as an exchange for reinforcing positions in the primary production, energetic or lower added value networks within the developing countries. The price of the deficit nourished on the expense of the enormous public debt will be later found equalized in the patrimonial balances of the most important transnational corporations.

The category of economies based on resources export comprises all those countries defined by preponderant export of energetic products as a main valence in creating the national payments balance of sheet. Out of the countries exporting energy, the most relevant ones are the producers and processors of oil and the exporters of natural gas. The economical improvements recorded by these countries depend on the demand curves recorded by the importing countries and, also, depend on the evolution of demand and offer formulated on the main international stock markets. The dependence is not unilateral, from the exporters of energetic resources to importers but it is a mutual relation, because the balance sheets of these economies are based mainly on the incomes cashed from this kind of exports. This implies a very high sensitivity in comparison to the production and consumptions' dynamic on a global level.

The rhythm of the economical growth and the balance sheet of payments for the oil exporters meet major fluctuations, much more significant than in the case of the importing economies. The balance of capital account of the net energetic products exporters can be seriously damaged on a short term, either by the massive withdrawal of investments in the exploitation and processing the energy or by the shrinkage of the worlds' production and the aggregate demand. This fact can be noticed by analyzing the decrease in terms of current balance account with almost 60% in 2006 compared to 2005, recorded by the exporters of oil despite the annual world growth rhythm of 1,3%, only based on the price stagnation and the cumulated reduction of demand from the UE's countries and the stagnation of consumption from the other developing countries. For 2008, the demand sluggish from the industrialized countries was possible to be compensated for a certain period only by increasing prices (on a record level), the trend practically reversing in oil exporters disadvantage starting with the end of September 2008. That was possible thanks to the reduction of demand from the main importers with over 36%. Therefore, the sudden fluctuations of the oil price on the international markets exist in the balance sheet of the states depending on the export of energetic resources with an immediate and multiplied effect.

Despite the strategic importance, these kinds of economies are considered vulnerable in their major dependence in relation with the demand conditions and investing capital cost. We can observe the exaggerated volatility of the 'black gold' price during to 2008. Within last year, we have recorded two price peaks, 34,6 USD/barrel (January 2009) as a minim compared to a price of 147,27 USD/barrel as a maxim. Meantime, the OPEC countries and its strategic allies did not succeed as easily as before to adjust the demand by manipulating the offer. The repeated restraints of production (up to 22% in only two weeks during to October 2008), only had a reverse effect, of reduction instead of increasing the price per barrel. Therefore, in crisis conditions, it is possible for the economies based on resources export to sluggish more deep owe to the 'lever' effect of the developed countries demand (the main consumers of oil and natural gas). To conclude, the basic characteristic of economies based on the export of energy it refers to the fact that follow the trend of world trade and it can contradict, during its evolution, the logic of resources rarity. Also, given the conditions of contracting production, these countries become more vulnerable.

Conclusions

The structure of oil demand has substantially changed during the past three years when the demand's structure, although quite constant as total volume, radically changed within its structure. This is due to the recorded requests of increase coming from the developing or least developed countries as the demand from the developed countries was decreasing. The calculated perspectives indicate a similar trend for the next two decades also, the increase being ascribed mostly to the developing countries demand. In this case, the sensitivity of the economies based on exporting energetic resources will be imperatively linked to the evolution of the world demand, but mainly to the macroeconomic situation of the developing nations with real chances to become the main upholders/supporters of the world consumption. Therefore, the volatility of the balance account of the 'energetic' countries shall depend on the production volume and on the engaged export by the developing, transition and least developed countries. On the other hand, following the logic of supply side economy theories, the economies based on export, will continue to depend on the aggregate demand from the developed countries as well. The most equilibrate solution for crises slowdown, benefic for all kind of economies shall be in this case only a general deflation. Taking under consideration the mentioned relations between these three kind of economics (or a mixed type as a particular perspective), is wrong to think that the credit stimulation will solve the problem of aggregate demand in developed countries. The global economical growth will be reinforced only on the base of value theory reappraisal.

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KEY PERFORMANCE INDICATORS USED IN TRANSPORT MODE BENCHMARKING

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Abstract: *Transportation modes are an essential component of transport systems since they are the means by which mobility is supported. Basically, transport modes are the means by which people and freight achieve mobility. They fall into one of three basic types, depending on over what surface they travel – land (road, rail and pipelines), water (shipping), and air. Each mode is characterized by a set of technical, operational and commercial characteristics. Each mode has its own requirements and features, and is adapted to serve the specific demands of freight and passenger traffic. This gives rise to marked differences in the ways the modes are deployed and utilized in different parts of the world.*

Keywords: *benchmarking, USA, industry, economic environment, KPI*

Introduced by the industry in the USA where first standards were formulated in 1987, **benchmarking** is the process of comparing the cost, time or quality of what one organization does against what another organization does. The result is often a business case for making changes in order to make improvements.

Benchmarking service performance in the area of freight transport across modes from the economic perspective of transport users can be extremely useful in nowadays economic environment, especially in the policy making process.

A comparison of performance between different modes requires analyzing key performance indicators (KPI) that could be clustered as quality indicators or belonging to different clusters such as:

- Economy
- Ecology
- Safety
- Security

1 Transport costs

The cluster costs include the following single performance indicators:

- Capital costs
- Fixed operation costs
- Variable operation costs
- Terminal handling cost

- Pre-haulage costs
- PricesLoad factor
- Fuel efficiency

Capital costs refer to equipment necessary for transport, such as ships, vehicles, wagons, containers etc. They are composed of repayment and interest.

Operating costs are partly fixed (labor, maintenance, repair, insurance or administration costs) and partly variable (fuel costs, port and terminal costs and charges). **Fixed costs** are unavoidable to keep the vehicle in an operational condition while **variable costs** depend on the route chosen and the distance. **Charges** can be defined as prices paid to an infrastructure owner (road, rail or inland waterway administration or investor).

A **terminal** can be described as a place with functions and technical assets to transship a load unit between two different modes of transport. Depending on the structure of the terminal the transfer could take place between various modes rail, road, sea and inland waterway. **The cost for terminal handling** can vary from region to region and in case of sea and inland waterway from port to port.

Pre-haulage costs arise when a load unit is hauled from the shipper to the terminal through a so-called pre-haulage leg, usually made by truck. In the terminal, the transship from the pre-haulage carrier to the main haulage, performed by the main non-road means of transport (e.g. short sea shipping or rail) takes place. In the area of intermodal transport the load unit is mostly hauled as well to a final consignee on an end haulage leg. Of course, more than one main haulage could exist just as more than one kind of terminal could be part of the intermodal transport chain. The pre- and end haulage to and from terminals is normally provided by road transport companies. These companies incur costs involved in the ownership and operation of vehicles, which in most cases include the payment of **taxes**. The total costs contain the time spent during loading and unloading as well as movement. Costs could also be incurred for the payment of infrastructure in form of **tolls**. For pre- and end-haulage often a **tariff per km** is applied which is, of course, much higher than the costs per km on the main haulage.

Prices can be described as the income received by an operator for a transport or terminal services; from the customer perspective the prices for such services become his costs.

The load factor represents the average use of capacity and is a factor for cost calculation per load unit. In some cases poor load factors can be generated by imbalance of flows in the two directions, seasonal pattern and strong day-to-day fluctuations. The load of the vehicle also plays an important role in the emissions of several pollutants. In addition, the load factor of a vehicle will also have an impact on fuel efficiency. The latter depends also on the type of engine and on the speed.

Dangerous goods require special treatment. Regulations for the transport are such detailed for the means of transport and specific for certain commodities that the shipper and specialized forwarder have to plan much more careful than for a normal transport. The calculation of costs is similar for general cargo but often a surcharge for dangerous cargo is requested. Such a surcharge can only be asked from the operator, e.g. the ferry operator. There are also restrictions to carry dangerous goods on ferries as long as passengers are on board. The capacity is limited and the stowage is a further restricting factor.

The indicators mentioned above can be summarized in the key performance indicator **“transport costs”** defined as total freight costs to the customer measured in costs per unit. The transport unit could be a container, semi-trailer or a tone of bulk cargo and the costs include the direct transportation charges plus any other costs associated with transportation paid by the user to move the transport unit from the point of commissioning to decommissioning.

The cluster includes a mixture of costs and prices. Costs are the better key figure because prices depend on the market and could include a major profit or a loss. Where prices are non-negotiable tariffs, the prices are suitable as key figures, e.g. in case of pre- and end-haulage. Where prices are not yet known because a new service is only planned, the calculation/estimation of costs is necessary. This is especially interesting for comparison of roro and lolo sea transport where costs also depend on ship sizes. If the potential transport volume is known it can be calculated if sea transport is less expensive then road transport.

Another case may also prove the necessity to know the costs and not only the prices. Ferry and roro shipping companies are free to calculate their prices. Many of them even avoid publishing their prices, only customers get information. A new customer will then, probably, pay the official price while good customers having many load units per year get high rebates. A difference of 20 or 30 % between official tariff and the price paid by such good customers can decide on the choice of this shipping company or route. In case of an empty return trip, when the ship may also be half empty, sometimes nearly every price is accepted by the ship operator. In any case, the position of the client who negotiates a price is better when he knows the costs of the operator.

2 External costs/socio-economic impact

The area of **external costs** and socio-economic impacts is very extensive. **External costs** relate to those costs which are incurred by other parties resulting from operator’s transport or terminal activities and some of them may also have socio-economic impacts. These are for example costs induced from accidents, air pollution, climate change and noise nuisance. An external cost is a cost that is not included in the market price, e.g. a cost that is not incurred by those who generate it. This means that when engaging in a transport activity, a person will incur private costs linked to the use of a mode of transport (tolls or fuel use), but will not be taking into account nuisances imposed on others such as congestion, accidents, noise, pollution and emissions of CO2.

In general, the Gross Domestic Product (GDP) of a country, the population and the level of unemployment can be seen as the main socio-economic indicators of a country’s development. Important socio-economic aspects are employment, education and training, working conditions, environment, quality of life, health and safety of the citizens. Each of the relevant transport modes can have effects on GDP or the social welfare, e.g. by the number of jobs created or wages paid. A comparison of the transport modes with regard to these social effects is difficult due to its complexity as well as in some cases the lack of information.

An example may be the evaluation of noise. Noise can be described as unwanted sound or sounds of duration, intensity or other quality that causes physiological or psychological harm to humans. Due to the complexity of noise, objective burdens are difficult to assess. The perception of sound as noise differs from person to person, from moment to moment.

For the cluster **external costs/socio-economic impacts** the following performance indicators have been identified:

- Infrastructure costs
- Air emissions
- Fuel consumption
- Fuel efficiency
- Noise emissions
- Injuries
- Deaths
- Employment
- Wages and salaries

Infrastructure costs cannot be completely taken into account due to the difficulty to differentiate between cargo transport and passenger transport. The same is true for **injuries** and **deaths**, because the relevant statistics only refer to the number of injuries and deaths by transport mode and do not distinguish between cargo and passenger transport.

Regarding **employment**, **wages and salaries** information is available about the number of employees directly involved in transport activities. The difficulty is the identification of the indirect impacts and, again, the distinction between cargo and passenger transport.

The KPI of this cluster can be summarized as the KPI "External costs", generally focusing on air and noise emissions, expressed in € per t km and defined as "costs to the public because of emissions of noxious gases". The fuel consumption is needed as information for emission's calculation.

3 Time

Time is not only relevant to the customer but also to the producers of the transport services, because it has a direct effect on costs. If transport assets move fast, the time-related costs of transport are reduced and transport efficiency is improved.

An important time factor in any logistics service is the period that elapses between the moment that demand becomes manifest and the moment of delivery. Logistics services must be able to deliver at the required time. The service does not just consist of moving the goods, but must also allow time for preparing them for dispatch, including operations such as order picking and handling, documentation and packaging, loading and unloading the cargo into the truck or loading unit, as well as transport as such.

In this cluster **time** can generally be defined as the total length of time between the point when the load unit is ready for transport and the point when it is delivered. The time for packaging, order picking, documentation and loading is not always taken into account because it depends on the capabilities of the shipper in its own premises and it should be similar for all modes. The possible performance indicators for this cluster are as follows:

- Speed of mode
- Transit time
- Terminal time
- Waiting time
- Frequency

The **speed of mode** defines the (average) speed of the transportation of goods. The average speed is influenced by the average speed on the various links (restriction of speed on roads, characteristics of tracks and waterways) as well as by delays at terminals, border crossing points and gauge transfers. Monitoring the **transit time** is more straightforward. For example in case of container transport it is the time span from the time of container availability for collection at the point of consolidation to the time of delivery at the point of de-consolidation.

With **terminal time** the average time in terminal including handling is meant. It describes the time a transport unit spends within the terminal. This involves the productivity of the terminal as well as the efficiency of the work. Furthermore, it includes the **waiting time** between entry and exit of the terminal. Another aspect of the waiting time refers, for example, to border crossing points as well as to frequency.

Frequency relates to how often a transport service is provided within a time period. The higher the frequency, the higher the quality of the service since it will more likely be available when desired. All of the above mentioned aspects will form the key performance indicator "Total transit time" measured in hours and referring to the average total time of regular service including transport, handling and waiting.

4 Reliability/punctuality

The cluster reliability/punctuality is composed of the following aspects:

- In time (punctuality)
- Condition of cargo
- Congestion
- Equipment breakdown

In general, reliability can be defined as the absence of unforeseen lowering of performance.

Transport services can be seen as reliable if goods are collected and delivered at the agreed time (**punctuality**) and arrive in good condition (**safety and security**).

Delays or errors always cause some inconvenience to the customer, whose departure and arrival schedules are disrupted. Sometimes logistics equipment and staff rosters have to be revised, which may reduce the efficiency of cargo handling operations.

Due to the fact that in practice it is not possible to guarantee a 100% punctuality there exists always some tolerance with respect to delays, in frequency and in time. To maintain high punctuality is more difficult when the time constraints are stricter. This is a trade-off customers as well as transport operators have to accept. Transport operators have to weigh this factor in the balance when scheduling their services: the promise of a fast transport system is more difficult to keep and any resulting loss of reliability will disappoint and possibly even alienate customers. Poor punctuality also generates costs for transport operators, because it has a negative effect on the use of transport equipment and drivers. Extreme delays can even affect subsequent assignments or require costly repositioning of vehicles and crew.

Punctuality in the area of transport can be influenced by a large number of factors. The more complex the form of transport service production, the greater the risk of delay and the more severe the constraints facing transport. One source of concern for punctuality is infrastructure capacity limitations, which cause congestion.

Congestion arises when traffic exceeds infrastructure capacity and the speed of traffic declines. It can be defined as a situation where traffic is slower than it would be if traffic flows were at low levels. The definition of these "low levels" (reference level) is complicated and varies from country to country.

This is most apparent in road transport in regions with dense traffic, but intermodal transport is also concerned. Infrastructure limitations can be overcome by extending capacity, by either construction or better management. The risk of delays may also be reduced by avoiding capacity constraints through rerouting- choosing other routes or other terminals- or by rescheduling, e.g. by opting for a different timing. A second type of risks relates to long transport processes, which are more exposed to disruption. This is the case with long distance transport and with (intermodal) transport chains with sequential phases and operations. Finally, the risk of delay increases with the number of players involved in the production of the transport service. The more players, the more room there is for imperfections in these agreements and in communications between the players. Even if there is no deviation from schedule within the chain, the different priorities of the different players may result in delays. Examples may be found in port terminals, where handling a vessel is postponed to cater to an ocean vessel arriving late in the port and in rail transport, where operators reassign drivers from freight to passenger trains.

Speed is also affected by **weather conditions** independent of the quality of the transport infrastructure. Longer periods of low temperatures may stop ship transport at all, especially on inland waterways. In this case reliable operators provide transport alternatives.

Rail-based transport chains, particularly in international traffic, are subject to all of the risks to punctuality described. Both Short sea and inland waterway traffic, for instance, have good images where their ability to deliver a punctual service is concerned. These modes use open infrastructure which offers the flexibility to adapt sailing speeds if it is necessary to catch up with sailing schedules. The risk of delay lies mainly in the seaports, where terminal handling capacity may be otherwise occupied. Most Short Sea and inland waterway services anticipate on this by allowing extra time margins.

All intermodal chains have a degree of flexibility to attenuate or completely offset delays. By timely and appropriate information about upcoming delays customers can also offset the consequences of late deliveries, which thereby increase their tolerance of delays.

The final key performance indicator relevant for this cluster is "**Delay**" expressed in average delay (hours). It is described as the average additional time to total transit time resulting from delays. In practice it would be the time elapsing beyond the agreed time window.

5 Accessibility/availability

For this cluster the following indicators can be extracted:

- Accessibility
- Access
- Terminal
- Network density
- Hours of operation

In general, **accessibility** can be seen as the ease with which the intermodal transport system can be used. **Availability** is very close connected to flexibility and is influenced by the same trends of more customer focus. It is not a very widespread indicator but there might be some possibilities of how to measure it and what should be included.

In the case of comparing different transport modes, the terms “**accessibility**” and “**availability**” rather refers to the transport connection as well as to the time required between the booking and the start of the transport. For example, in maritime transport the depth of water available in a port will constrain the maximum size of **vessel** that can be deployed. By far the most common cause of delays to shipping in the container transport industry is the lack of berth availability. If the vessel cannot come to berth on schedule, the implications for the efficiency, timeliness and predictability transport movements are obvious. Such consequences of availability are already included in the KPI **Delay**.

Terminals can usually be seen as transfer points that are the principal component of an intermodal transport chain, constituting the node where transshipment of goods from one mode to the other takes place. The main terminal types are continental terminals, inland waterway terminals and maritime terminals. The major difference between maritime terminals and inland (continental and inland waterways) terminals is the size and the types of loading units (mainly in Europe). Volumes handled at seaport terminals are much higher than in inland terminals.

The **network density** is a more theoretical figure and refers to the physical infrastructure. It can be described in terms of number of network elements and total kilometers of the links.

Obviously, this is the total transport infrastructure, regardless of whether intermodal transport uses it. If a terminal can be used depends on the individual location and services offered, not on a theoretical figure like network density etc.

The key performance indicator for this cluster will be “**Availability**”, defined as the minimum time required between booking and start of transport measured in hours.

Where no common carriers with fixed schedules are used but individual transport solutions the chartering of a ship or special vehicle and its positioning in the port of departure adds to that time.

6 Flexibility/organization

Flexibility describes the ease with which the transport system adjusts to an unexpected change in the logistic requirements, whereas organization will have a major impact to flexibility. Flexibility can be seen as the transport mode's adaptability to changes in the chain both on a short time basis and on a longer time scale. It can be the possibility to handle specific customer orders within, and without disturbing, the normal transport flow.

In this cluster there are indicators which are of a more qualitative nature. One has to measure one thing and then use that to indicate something that is basically not measurable. The following indicators and headwords have been extracted for this cluster:

- Responsibility
- Number of partners
- EDI (Electronic Data Interchange)
- Compatibility of EDI
- Compatibility of load units
- Documentation
- Tracking and tracing
- Information accessibility
- Information on deviation
- Ease of use

Flexible logistics services are services which are capable of coping with unforeseen fluctuations in demand or circumstances. **Flexibility** implies responsiveness. No time should be lost in mobilizing capacity - assets and labor - to handle and transport the goods. When transport is vulnerable to extreme weather conditions, infrastructure blockages or congestion, logistics service production requires internal flexibility to cope with changing circumstances.

Concerning the number of partners one can state that the more partners are involved in a transport flow the more flexibility is decreasing, because there are too many responsibilities.

Flexibility will be optimal when only one partner – e.g. a truck operator – is involved.

The **compatibility** of load units in general cannot be seen as optimal. For example, ISO containers are not optimized for European pallets. A truck/trailer or a pallet-wide container contains more pallets than a 40' container. Railway containers are not stackable on ships, railway gauges and profiles differ from country to country etc.

7 Safety and security

Safety and security themselves are quality requirements. Because of public interest in safety, there are many regulations and strict enforcement of the provisions governing the technical state of the equipment, staff qualifications and working conditions, such as driving hours. On top of this, shippers and transport operators do much to increase safety on a voluntary basis. High risks and the heavy impact of incidents will immediately be reflected in insurance costs.

Safety requirements may be expressed in terms of the degree of acceptance of risks to third parties or the environment. High-impact accidents, e.g. with personal injuries or with major damage to the environment, are rare in any of the freight transport modes.

The small number of such accidents makes it complicated to make a proper assessment of logistics safety levels. Safety comprises internal factors, such as personnel skills and qualifications, but is also subject to external factors like traffic safety. The degree of reliability can only be measured once the service has been performed: logistics services are deemed reliable if experience shows that promises are kept."

Security relates to the level of security on the transport system and at interchanges. It is assessed on a qualitative basis. It relates more to missing or violated (not destroyed by accident) cargo, for example by theft or terrorism.

In this cluster the following indicators are found:

- Accidents
- Damage
- Loss of Cargo
- Fire
- Dangerous cargo
- Theft
- Terrorism

Accidents can on the one hand result in injuries or death of persons and on the other hand in damage or loss of the cargo or parts of it as well as to delays in delivery. While the number of accidents is different according to the modes, the decisive issue here is what happens to the cargo.

Damage to cargo leads to a loss in its value, which may exceed the full intrinsic value of the shipment. Damage can be caused by improper handling during logistics operations (contamination, breakages, incompleteness or insufficient packaging, e.g. perishable cargo) or by external factors such as traffic accidents.

Theft and fire can be reasons for the damage or loss of cargo whereas terrorist attacks are more often aimed to destroy human lives or transport equipment and to a lesser extent the cargoes themselves.

In a reaction to **terrorist attacks** the IMO has released the new ISPS Code requiring security plans and security officers in shipping companies, on board and in ports. Access to terminals for international sea transport has been restricted by checkpoints and fences. Ports have decided to levy a security fee per load unit to cover the costs for security. Typical fees per container are 5 to 10 Euros.

For this cluster the KPI "**Safety**" is of utmost relevance. It is defined as the risk of financial damage that can be mitigated by paying insurance premiums and will be measured in € per load unit. Regarding security, additional fees are levied, which can be directly used as key figures.

8 Political/regulatory issues

There are various differences in national policies among EU member states (and nonmember states), well-known differences and performance unevenness in physical infrastructure throughout Europe due to historical factors and conditions such as underinvestment or over-usage (congestion), and various other natural (e.g. distance) and artificial (e.g. local subsidies) influences on logistics systems.

The number of regulations to follow is dependent on the number of modes used and the number of countries involved in the transport chain. This cluster refers to more qualitative indicators which are:

- Restrictions
- Technical standards
- Harmonization
- Liberalization
- Privatization
- IMO Conventions
- Dangerous goods

Regarding **restrictions** these could be the limitation for road transport operations, e.g. the maximum gross mass allowed for road transport, driving bans, labor laws for road operations: the payment of all working periods, speed restrictions or the limitations for terminal operations: hours of operation, handling of dangerous cargo.

Technical standards define the minimum requirements of infrastructure and transport equipment.

All these indicators described are difficult to measure, a basis could be the number of regulations to follow, but as already mentioned this depends on the number of modes and countries involved in the transport chain. Therefore the KPI "Regulations" – defining the framework conditions – will be assessed by a ranking, whereas the most positive ranking will be assigned to the lowest number of regulations.

Benchmarking is a useful means of analyzing and comparing different transportation modes in an attempt to find out the best combination in terms of economic efficiency, while observing environment and safety regulations. Benchmarking is increasingly recognized as a powerful tool when working in the ever shifting field of world economy.

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ECONOMIC ACRONYMS

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Abstract: *Abbreviations are known as a shortened form of a written word or phrase used in place of the whole. They come in many forms and are extensively used in all fields of activity, and the economic language is no exception. It abounds in all sorts of short forms that make communication one hand easier for those who are familiar with it, and on the other hand quite frustrating for outsiders of the field. Economic acronyms cover all sorts of economic realities, from ordinary concepts (IR – interest rate), to currencies, to economic events in the life of a company (AGM – Annual general meeting), to complex terms (CES – constant elasticity of substitution, a property of production) and intricate economic models (IS-LM).*

Keywords: *abbreviations, interest rate, annual general meeting, constant elasticity of substitions, initialism*

An abbreviation – from the Latin word *brevis* meaning "short" – is a shortened form of a word or phrase. Usually, but not always, it consists of a letter or group of letters taken from the word or phrase.

There are several types of abbreviations, the most common being acronyms and intialisms, syllabic abbreviation, truncations, portmanteaus.

An acronym is a word formed from the initial parts (letters or syllables or arbitrary parts) of a name. One of the most well-known examples is NATO (North Atlantic Treaty Organization). Other acronyms are formed using parts of words, as in Benelux.

The word acronym was coined in 1943 for abbreviations pronounced as words, such as NATO, AIDS, Laser or SARS (Severe Acute Respiratory Syndrome).

An initialism is a group of initial letters used as an abbreviation for a name or expression, each letter being pronounced separately, for example, "BBC" (British Broadcasting Corporation), or "PBS" (Public Broadcasting System). An initialism originally described abbreviations formed from initials, without reference to pronunciation.

The key difference between an acronym and an initialism is that an acronym forms a new word, while an initalism does not. For example, you say "nay-to" for NATO; this means you are saying a word, as opposed to saying each letter (ehn-ay-tee-oh). So "NATO" is an acronym. But "U.S.A." is an intialism for United States of America - you say each letter individually (you don't say "yusa", so you know it's not a word).

Also, the periods used when writing it are a clear indication that it is an intialism.

There are other types of abbreviations as well. Consider, for example, the truncation, which is an abbreviation of a word consisting only of the first part of the word. Most often used in a context (such as for mail) where certain words must be written (and read) repetitively. Examples: Tues. = Tuesday; Dec. = December; Minn. = Minnesota; Eur = Europe, European.

Also consider syllabic abbreviations – abbreviations formed from (usually) initial syllables of several words, such as Interpol = International + police, INMARSAT (INternational MARitime SATellite) or Incoterms (International commercial terms).

Syllabic abbreviations should be distinguished from portmanteaus. A portmanteau is used broadly to mean a blend of two (or more) words, a word formed by blending sounds from two or more distinct words and combining their meanings. Some examples are brunch (breakfast + lunch), Tanzania (Tanganyika and Zanzibar), spork (an eating utensil that is a combination of a spoon and fork) or chocoholic (a person who is addicted to chocolate).

Of all the above-mentioned names, acronym is the most frequently used and known; it is widely used to describe any abbreviation formed from initial letters.

The complex language of economics uses many such short forms for different economic concepts but especially to refer to different institutions and their programmes.

Some of the most common economic terms are used in the form of acronyms. There are acronyms used to identify taxes and tariffs, rates and indexes, quotas and quotations, or various measures.

GDP (Gross Domestic Product) and GNP (Gross National Product) are main measures of national economic activity.

ACT, CGT, CTT, CET, and VAT – they all stand for different types of taxes or tariffs – ACT (Advance Corporation Tax), CGT (capital gain tax), CTT (capital transfer tax), CET (common external tariff), VAT (Value Added Tax). CPI and RPI are indexes, standing for consumer price index and Retail Price Index. LIBOR reads London Inter Bank Offered Rate, MLR stands for minimum lending rate, IR for interest rate, NAIRU is the non-accelerating inflation rate of unemployment and NAWRU non-accelerating wage rate of unemployment. Even though it does not have the word rate in its name, MEC (marginal efficiency of capital) is also a rate, the highest interest rate at which a project could be expected to break even.

Earnings, return and capital are three examples of basic economic terms. We can find them in the following acronyms: EBIT (earnings before interest and taxes), EPS (earnings per share), P/E ratio (price to earnings ratio) and ROCE (return on capital employed), ROIC (return on invested capital).

Small businesses (SB) or, in other words, small and medium enterprises (SMEs) play a crucial role in the economic network. Therefore, we can come across many acronyms containing the letters SB: SMA (Small Business Administration), SBDC (Small Business Development Center), SBI (Small Business Institute), SBIR (Small Business Innovation Research).

Economic models are either named after the economist(s) who built them: the Harrod -Domar model, Heckscher-Ohlin model, or by means of an acronym: the IS-LM model, CAPM (capital asset pricing model), CGEM (computable general equilibrium model).

Some of the economic mechanisms are known as an acronym, such as ERM – Exchange Rate Mechanism, as well as some economic methods of analysis – RAROK – risk-adjusted return on capital, a method of comparing returns on different investment taking account of risk.

Economists refer to various schemes, plans or strategies using acronyms: BES means Business Expansion Scheme, ESOP and PEP mean employee stock ownership and personal equity plan, whereas MTFS stands for Medium Term Financial Strategy.

R&D (short for Research and Development) is an economic activity involving the use of resources to create new knowledge and to develop new and improved products or more economic methods of production.

RPM, TESSA, VER are economic practices. The once used RPM (resale price maintenance) was the fixing by manufacturers of minimum prices at which their products could be resold by distributors. TESSA (tax-exempt special savings account) in the UK indicates that individuals can invest a limited amount each year with a building society, the interest being tax-free. VER (voluntary export restraint) identifies an agreement by a country's exporters or government to limit their exports to some other country.

DCE and LBO are only two examples of economic phenomena. DCE (domestic credit expansion) is an increase in the money supply not due to balance-of-payments (BOP) surplus. LBO (leveraged buy-out) depicts a change in control of a company financed by borrowings.

Many economic acronyms refer to the world of money: SDR (Special Drawing Right) form of international money created by IMF (International Monetary Fund), USM (Unlisted Securities Market), FDI (foreign direct investment), IR (interest money), but others refer to realities that have nothing to do with money whatsoever: NTB (non-tariff barriers) man-made obstructions to international trade other than tariffs.

The largest category of economic acronyms groups institutions (agencies, associations, authorities, bureaus, boards, banks, commissions, councils, corporations, confederations, departments, federations, funds, institutes, offices, organizations, services, systems, unions) and their programmes (agreement, convention, conference, policies). By their names, some of these can be easily recognized as being British, while others are obviously American institutions, many being, of course, regional or international.

Banks are important players in the economic world. The bigger they are, the more important. BIS (Bank of International Settlements), EBRD (European Bank for Reconstruction and Development), ECB (European Central Bank), IBRD (International Bank for Reconstruction and Development) are among the most influential of them all.

US Economic Acronyms

IRS (Internal Revenue Service) is probably the best known acronym for an American institution, being charged with assessing and collecting personal and business federal taxes.

Another well known agency in the United States with wide environmental and economic implications is the EPA (Environmental Protection Agency).

The US Department of Commerce which assembles and publishes the US national income accounts is the BEA (Bureau of Economic Analysis). Another example of bureau is the NBEAR (National Bureau of Economic Research), a major US provider of high quality analysis of micro, macro, and international aspects of the US economy.

EEOC (Equal Employment Opportunity Commission) is a US federal commission which deals with discrimination in wages, hiring, firing, promotion, and training on the basis of age, color, race, sex, religion, or national origin.

FTC (Federal Trade Commission), FSLIC (Federal Savings and Loan Insurance Corporation), ICC (Interstate Commerce Commission), FCO (Federal Cartel Office), by their mere names, send to their US origin. Other examples of commissions are MMC (Monopolies and Mergers Commission) – a UK body which investigates anti-competitive practices and SEC (Securities and Exchange Commission) – a US government agency which supervises trading in securities and takeovers.

Two examples of councils are the British ESRC (Economic and Social Research Council) and NEDC (National Economic Development Council).

CCC (Commodity Credit Corporation) is a US federal agency set up to provide price support for US farmers. FSLIC (Federal Savings and Loan Insurance Corporation) is another US institution, while IFC (International Finance Corporation) is in fact an international investment bank affiliated to the IBRD or World Bank.

UK Economic Acronyms

SIB (Securities and Investment Board) is a regulatory body set up to oversee UK financial markets. PIA (Personal Investment Authority) is a British organization regulating the branches of the investment business dealing mainly with private investors. CBI (Confederation of British Industry) is a federation of UK companies, mainly from the manufacturing sector. NIERS (National Institute of Economic and social Research) is an independent UK body carrying out research into both macro- and microeconomic aspects of economy. CSO (Central Statistical Office) publishes many UK statistical sources. OFT (Office of Fair Trading) administers UK competition policy. ECGD (Export Credits Guarantee Department) is responsible for encouraging UK exports by insuring exporters against risks.

European and Other Regional Economic Acronyms

EBRD (European Bank for Reconstruction and Development), ECB (European Central Bank), EMI (European Monetary Institute), EFTA (European Free Trade Association), EMS (European Monetary System), EMU (European Monetary Union) are only a couple of economic acronyms belonging to the new European reality – the EU. Another example is the CAP. CAP stands for Common Agriculture Policy. Just as it name shows, it is an EU policy towards agriculture.

NAFTA, LAFTA, EFTA, CEFTA are regional free trade agreements for North America, Latin America, Europe, and Central Europe, respectively. For the Asian continent there is ASEAN (Association of Southeast Asian Nations).

International Economic Acronyms

A frequently mentioned international organization is the OPEC (Organization of Petroleum Exporting Countries) which co-ordinates the policies of oil-producing countries in negotiating with oil companies.

IMF (International Monetary Fund), UNCTAD (United Nations Conference on Trade and Development), CITES (Convention on International Trade on Endangered Species), ILO (International Labor Organization) are other examples of international organizations known and pronounced as acronyms.

FAO (Food and Agriculture Organization) is an agency in the UN responsible for problems of agricultural production and nutrition. OECD is an international organization set up to assist member states to develop economic and social policies to promote sustained economic growth with financial stability.

UNDP (United Nations Development Programme) is a body which gives technical assistance and makes soft loans to LDCs (less developed countries).

Incoterms or international commercial terms are a series of international sales terms widely used throughout the world. They are used to divide transaction costs and responsibilities between buyer and seller and reflect state-of-the-art transportation practices. They include: EXW (Ex Works), FCA (Free Carrier, FAS (Free Alongside Ship), FOB (Free On Board), CFR (Cost and Freight), CIF (Cost, Insurance and Freight), CPT (Carriage Paid To), CIP (Carriage and Insurance Paid to), DAF (Delivered At Frontier), DES (Delivered Ex Ship), DEQ (Delivered Ex Quay), DDU (Delivered Duty Unpaid), DDP (Delivered Duty Paid).

Knowing the meaning of all these acronyms is essential for an economist's everyday activity. Economic literature is abounds in abbreviations of all sorts, this being one main feature of the economic language.

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NEGOTIATION AND PROCESS SYLLABUS IN PRACTICE

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Abstract: *The present paper is based on the assumption that collaborative classroom negotiation could be introduced into an ESP teaching situation where previously design and implementation decisions had been made prior to the start of the course. A process syllabus evolves directly from the immediate context of the course in which it is developed, deciding who does what as defined by whom. Any coherent language curriculum will attempt to reconcile what is desirable [policy] with what is acceptable and possible [pragmatics]. This paper will attempt to demonstrate how classroom decision-making could provide a pivotal role in facilitating this reconciliation by means of negotiating learning content.*

Keywords: *syllabus, collaborative learning, learner centredness, autonomy*

1 Relating negotiation to autonomy and the process syllabus

The key concepts of negotiation in the classroom decision-making according to Breen and Littlejohn's view (2000) seem to be closely interrelated with the complex notions of collaborative learning, learner centredness and autonomy. The authors of "Classroom Decision-Making" proceed to explore the nature of classroom negotiation on three levels: i) *personal* negotiation, deeply rooted in one's psychological processes and involving complex mental processes; ii) *interactive* negotiation, based on research into the nature of conversational interaction and having significant implications for language learning; and iii) *procedural* negotiation which is viewed as the closest type of negotiation to various real life social contexts.

All these levels echo Benson's categories of learner autonomy at least as far as "psychological" and "political" autonomy types are concerned (1997: 29) in that they involve an internal cognitive and deconditioning process on the part of the learner for the former type, and a recognition of the rights of learners within educational systems to negotiate external curricular requirements for the latter type. Additionally Breen and Littljohn's definitions of interactive and procedural negotiation allude to Kohonen's conception of autonomy (1992) as taking place within an interdependent classroom with importance being placed on each individual to accept responsibility for their own conduct to create a supportive environment encouraging work towards shared goals and their solving of potential conflicts in constructive ways.

The concept of the classroom as a "social context" included in the definitions of the interactive and procedural definitions of negotiation reflects Breen and Candlin's observation (1980) of the current classroom shift in focus into the domain of the power relationships between teachers and learners. Thus the book under analysis appears to point a way towards the achievement of greater democratisation of educational systems, a trend which has come to be perceived as a core element in the educational policy making process (Girard et al, 1994; Stoks, 1996). Similar to the case of autonomy in learning, the process of negotiation in the classroom is a highly debatable subject matter which could be viewed as either a precondition for effective learning, or as a mere "distraction from the real business of teaching and learning languages" (Benson, 2001: 1).

Breen and Littljohn's position within the ongoing debate on whether current curricula could be considered effective and valid from learners' needs and agendas point of view, is embedded in their strong belief that the learners have the right to make choices concerning their learning, assuming a proactive role on various levels, such as the learning management or the learning content, drawing on previous experience "as a means of shaping future action through informed choice" (Breen, 2000: 283). Therefore they advocate the implementation of a process syllabus which provides an infrastructure rather than a learning plan, with the syllabus designer no longer pre-selecting learning content, but providing a framework for teacher and learners on which to create their own ongoing syllabus in the classroom through procedural negotiation in order to reach "a shared understanding" of external requirements and learners' individual agendas (ibid: 9). This can allow for changing abilities, learning needs and perceptions in the learners, without specifying particular content, methodology, lexis or grammar. Thus a process syllabus enables both parties to share responsibility at the preplanning, teaching and evaluation stages of the planning process through a cyclic approach to negotiation. This procedure is aimed at encouraging learners' ongoing reflection on ways in which they believe that their learning needs could be met.

Breen and Littlejohn's discussion of the primary theoretical and empirical origins of procedural negotiation comprises four major perspectives which allude once again to the concept of autonomy in the learning process. Their views refer to both learner and language learning:

1. A view of the learner as an active agent of his or her learning. This aspect clearly relates to Rathbone's definition of an autonomous learner as: "a self-activated maker of meaning, an active agent in his own learning process. He is not one to whom things merely happen; he is the one who, by his own volition, causes things to happen. Learning is seen as the result of his own self-initiated interaction with the world"(in Thanasoulas, 2000). The proactive learner will therefore possess a certain attitude to learning i.e. one in which he/she is prepared to assume more responsibility for their own learning, assuming a more active and independent involvement with the target language.

2. A view of learning as emancipatory in contrast to conventional education. This idea seems to derive directly from Holec's report for the Council of Europe in which he states that adult education "becomes an instrument for arousing an increasing sense of awareness and liberation in man, and, in some cases, an instrument for changing the environment itself"(in Little, 1991: 6).

3. A view of learning as contextualised in a wider society in which student responsibility and co-operation can be seen as expressing and enabling participation as a citizen in the democratic process. Once again the idea of autonomy in learning with direct

implications into a wider societal context pervades the above statement in that education is capable of removing “the barriers between learning and living” (ibid: 8).

4. A conception of language learning as seen in social and cultural action wherein what is learned and how it is learned are collaboratively shaped.

The basic proposition underpinning the issues presented, including the issues of autonomy, is essentially that “language reflects the contexts in which it is used and the purposes to which it is put”, (Nunan, 1991: 124). The notion of interaction entails communication and is based on three processes: interpretation, expression and negotiation of meaning which are intrinsically related to language learning.

The primary objective is to enhance the users’ ability to participate in these situations (Widdowson, 1978) through awareness of the various factors or competencies that are employed in order to achieve effective interaction (Canale and Swain, 1980). This leads to their creative utilisation and an ability to negotiate the rules and conventions of communication in practice (Breen, 1984). Therefore the notion of negotiation is clearly connected with the notion of interaction in the form of syllabus negotiation which becomes synonymous with the learning process itself. By communicating within the context of classroom interaction about the curriculum and the learning process, the learners will gradually shift from negotiation of meaning to negotiation as a process of discussion to reach agreement drawing closer to the process of negotiation for the ‘real’ world.

Furthermore it is claimed that negotiation could provide support for the learners’ ability to reflect on and articulate ways in which their perceived needs might be more accurately met as well as enhance their ability to actively engage in the process of life-long learning. These ideas once again are related to the notion of autonomy in learning embraced by progressive theoreticians like Holec (1981), Clark (1987) and Little (1991) who perceive learners as active agents of change within the society context. Thus education represents an empowering factor enabling people to take charge of their future: “producers of their society” (in Little, 1991: 6).

2 A review of classroom applications of the process syllabus

In terms of practical application of negotiation in the classroom, initial attempts could prove challenging especially when the teacher does not share the background of the students as in the case of Nikolov who found out that age is a major impediment to negotiation with primary age schoolchildren in Hungary (in Breen, 2000: 85). Another difficult case reports that highly specialist content of an ESP syllabus proved to be a daunting task for the teacher. Nevertheless the adoption of a negotiated syllabus seemed to be a viable solution in order to enable Martyn to enter the culture of nursing and consequently to more accurately address the students’ needs for ESP (ibid: 154).

However I would argue that in spite of the fact that negotiation may prove useful in the case of ESP content, it is a rather time-consuming process. Although Cook’s comments on the issue seem encouraging underlining that the time spent negotiating decisions is “not so much a case of requiring more time as a case of reorganising available time”(1992: 27), there is still a need for further practical evidence. Slembrouck’s Belgian university context (in Breen, 2000: 139) attempts to address the problem of constraints by making use of the teacher’s determined choice of syllabus content in an arbitrary way. The teacher consulted the students about their learning priorities in an overt manner through interactive negotiation and reached agreement on syllabus content. The procedure appears quite simplistic and easy to implement, however practical work may reveal that certain groups of students simply cannot reach a consensus every time negotiation takes place thus impeding the development of the very core notion of negotiation i.e. reaching a shared understanding of both external requirements and different learning needs.

Norris and Spencer report another case of successful implementation of negotiation and classroom decision-making (in Breen, 2000: 195). They present an English language course for vocational teachers and managers from Indonesian colleges where obviously there is a heterogeneous student body who share common interests and collaborate in order to facilitate learning. Here the benefits of negotiation are very clear. However there is one question that may arise when analyzing the present case: would the outcome of negotiation be equally successful if the participants had not had any previous experience of negotiation techniques either in terms of their professional life or past learning experiences?

Without defying logic, one may assume that successful negotiation and implementation of a process syllabus could be done almost exclusively with adult learners who possess a certain degree of sophistication both in terms of professional and formal educational experience. Therefore teacher-contributors to the present volume besides primary or secondary school students included in the process of negotiation stakeholders, parents and/or other teachers. Presumably such course of action involving negotiation on a larger scale which extends beyond the classroom boundaries might add validity to their research. In spite of the efforts of teacher-contributors, certain case studies appear rather unrealistic and inconclusive (see “Negotiating the syllabus: learning needs through pictures” by Edmundson and Fitzpatrick in Breen, 2000: 163).

As can be seen from all aforementioned cases of classroom decision-making there are certain limitations to Breen and Littljohn’s theoretical arguments. One is that the process of negotiation could not be carried out unless certain conditions are met:

1. The learners should have some previous experience with regard to formal education considering they are to be involved in decision making sharing based on *informed* choices.

2. The learners should possess some experience of metalanguage required when discussing their choices.

3. The learners should be accustomed to negotiation procedure.

4. The learners should display an open-minded and flexible attitude and aim to avoid cultural barriers.

All these preconditions for negotiation appear as highly demanding for both teacher and learners considering the specificity of various learning settings, the learners’ diverse psychological characteristics and different learning styles, the appropriateness of negotiation set within certain cultural contexts, time constraints, the size of the groups of learners et cetera. Ivanic suggests a possible way to resolve the tension between students’ engagement in negotiation and the fact that they may not be very good at negotiating through the introduction of an induction phase, aimed at giving students the tools to benefit from the negotiation process (in Breen, 2000: 246).Nevertheless the other preconditions for negotiation have not been considered and the question of whether they could be met or not is still to be investigated.

3 Conclusion

Various contributors to Breen and Littlejohn's volume tried to demonstrate that most of the aforementioned constraints could be overcome in various teaching setting and classroom contexts in order to arrive at a successful implementation of the process syllabus in action. However many of them come across as lacking substance and practical applicability due to the very constraints that they had tried to resolve.

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THE CONSCRIPTION IN FRANCE

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Abstract: *Although it is already known that France discarded conscription almost 8 years ago, this paper attempts to investigate the reasons for the end of conscription in the above-mentioned country, after a brief presentation of the French military and its organisation, a history of the word itself, and last but not least, the arguments for and against this ancient institution strictly linked to the republican values.*

Keywords*: conscription, France, draftees, military, JAPD*

The Military of France encompasses an army, a navy, an air force and a military police force. The French armed forces are divided into four branches: 1. ***Armée de Terre*** - Army, including: *Infanterie* - Infantry, Chasseurs Alpins - mountain infantry, *Arme Blindée Cavalerie* - Armoured Cavalry, *Artillerie* - Artillery, Foreign Legion (infantry, cavalry, engineers), Troupes de marine - Marines (infantry, cavalry, paratroopers, artillery), *Aviation Légére de l'Armée de Terre* - *ALAT* - Army Light Aviation, *Génie* - Engineers including the Paris Fire Brigade, *Transmissions*- Signals, *Train* - Transport and logistics, *Matériel* - Supply; 2. ***Marine Nationale*** -Navy, including: Naval Air, naval fusiliers (naval ground troops) and naval commandos including the Marseille Fire Battalion; 3. ***Armée de l'Air*** - Air Force including territorial Air Defence, air fusiliers (air force ground troops); 4. ***Gendarmerie Nationale*** - Gendarmerie, a military police force which serves for the most part as a rural and general purpose police force. In 2009 this force will quit the ministry of defence to fully join the ministry of interior (police).

The titular head of the French armed forces is the President of the Republic, in his role as *Chef des Armées* - the President is thus Commander-in-Chief of French forces. However, the Constitution puts civil and military government forces at the disposal of the *government* (the executive cabinet of ministers, who are not necessarily of the same political side as the president). The Minister of Defence (as of 2007, Hervé Morin) oversees the military's funding, procurement and operations.

Historically, France relied a great deal on conscription to provide manpower for its military, in addition to a minority of professional career soldiers. Following the Algerian War, the use of non-volunteer draftees in foreign operations was ended; if their unit was called up for duty in war zones, draftees were offered the choice between requesting a transfer to another unit or volunteering for the active mission. In 1996, President Jacques Chirac's government announced the end of conscription and in 2001, conscription formally was ended. Young people must still, however, register for possible conscription (should the situation call for it). A recent change is that women must now register as well.

France has undertaken a major restructuring to develop a professional military that will be smaller, more rapidly deployable, and better tailored for operations outside of mainland France. Key elements of the restructuring include: reducing personnel, bases and headquarters, and rationalistion of equipment and the armaments industry.

On 31 July 2007, President Nicolas Sarkozy ordered M. Jean-Claude Mallet, a member of the Council of State, to head up a thirty-five member commission charged with a wide-ranging review of French defence. The commission issued its White Paper in early 2008.[5]. Acting upon its recommendations, President Sarkozy began making radical changes in French defence policy and structures starting in the summer of 2008.

Conscription (also known as **Compulsory Service**, **The Draft**, the **Call-up** or **National Service**) is a general term for involuntary labor demanded by an established authority. It is most often used in the specific sense of government policies that require citizens to serve in the armed forces.

The term “conscription” refers only to the mandatory service; thus, those undergoing conscription are known as *conscrit* “conscripts” or “selectee”. In France, military age is 17. Since the Algerian War of Independence, conscription has been steadily reduced and was abolished by the government of Jacques Chirac in 1996.

Modern conscription was invented during the French Revolution, allowing the Republic to defend itself from European monarchies’ attacks. Deputy Jean-Baptiste Jourdan gave its name to the September 5, 1798 Act, whose first article stated: *“Tout Français est soldat et se doit à la défense de la patrie”*, that is ”*Any Frenchman is a soldier and owes himself to the defense of the nation.”* However, the compulsory National Service is actually suspended (law October 97) for anyone born after December 31st, 1978. The actual Army obligations of any person (male and female alike) is to get registered, and to spend a day -*Journée d'appel de préparation à la défense* (JAPD) - to be ”selected”. People not doing so cannot take a national exam (including driving license) until 25 but this ”punishment” is cancelled as soon as they submit.

What are the arguments for and against the draft? The issue is a classic debate between individual liberty and duty to society. Democracies value individual liberty and choice; however, democracy does not come without costs. How should those costs be shared?

It is a fact that warfare has dramatically changed since Napoleon’s march to Russia or the battle of Normandy. It has also changed since Vietnam. There is no longer a need for massive human cannon fodder, with the military going ”high tech” and all. Thus one major **argument** against the draft makes the case that highly skilled professionals are needed, not just men with combat skills.

As far as the pros are concerned, we would like to quote Charles Moskos, a former draftee who was of the opinion that a draft would dramatically upgrade the quality of recruits, because it would give the military access to a true cross-section of our youth.

All in all, conscription was abolished in France in 2001, June 27, and it seems to be the end of conscription in Europe.

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WHAT ABOUT HOTEL CALIFORNIA?

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***Abstract**: Considering the ever-changing preferences of worldwide tourists we have deemed proper to shed some light upon the roots of the word hotel as it prevails in the English language in particular and the cross-cultural context of nowadays society. In the paper you shall find some interesting approaches in terms of usage and connotations of the word hotel as well as a list of newly-formed words in the field of hotel management and administration. Last but not least we have included some hints at a renowned hit namely Hotel California of the American hit The Eagles all in the hope of focusing on this “Ba’hai” word in today’s globalized planet.*

***Keywords**: hotel, etymology, connotation, denontation.*

During the Middle Ages the hospital could serve other functions, such as almshouse for the poor, hostel for pilgrims, or hospital school. The name comes from Latin hospes (host), which is also the root for the English words hotel, hostel, and hospitality. The modern word hotel derives from the French word hostel, which featured a silent s, which was eventually removed from the word. (The circumflex on modern French hôtel hints at the vanished s.)

Grammar of the word differs slightly depending on the dialect. In the U.S., hospital usually requires an article; in Britain and elsewhere, the word is normally used without an article when it is the object of a preposition and when referring to a patient ("in/to the hospital" vs. "in/to hospital"); in Canada, both usages are found.

The definition of the word as it appears in the dictionary is as follows:

– noun

1. a commercial establishment offering lodging to travelers and sometimes to permanent residents, and often having restaurants, meeting rooms, stores, etc., that are available to the general public.

2. (initial capital letter) Military. the NATO name for a class of nuclear-powered Soviet ballistic missile submarine armed with up to six single-warhead missiles.

3. a word used in communications to represent the letter H.

[Origin: 1635–45; < F hôtel, OF hostel hostel]

— Synonyms 1. hostelry, hostel, guesthouse, motel. Hotel, house, inn, tavern refer to establishments for the lodging or entertainment of travelers and others. Hotel is the common word, suggesting a more or less commodious establishment with up-to-date appointments, although this is not necessarily true: the best hotel in the city; a cheap hotel near the docks.

The word house is often used in the name of a particular hotel, the connotation being wealth and luxury: the Parker House; the Palmer House. Inn suggests a place of homelike comfort and old-time appearance or ways; it is used for quaint or archaic effect in the names of some public houses and hotels in the U.S.: the Pickwick Inn; the Wayside Inn. A tavern, like the English public house, is a house where liquor is sold for drinking on the premises; until recently it was archaic or dialectal in the U.S., but has been revived to substitute for saloon, which had unfavorable connotations: Taverns are required to close by two o'clock in the morning. The word has also been used in the sense of inn, esp. in New England, ever since Colonial days: Wiggins Tavern.

1. "The origin of the word hotel."

The Latin word hospes has given rise to a whole family of related words. Its primary meaning was "host" (which is its English descendant), but it could also mean "guest" (those wacky Romans, eh?). Hospes was the source of the Latin word hospitalis "hospitable" and of the Late Latin hospitale "large house or inn" which gave English the word hospital, first recorded in 1300. In Old French, hospitale was simplified to hostel, which entered Middle English as both hostel and ostel and became the Modern French hôtel, entering English as hotel in the early 18th century. The word ostler, meaning "one who tends to horses", is thus essentially the same word as hotelier, the difference being that ostler dates from 1376 and hotelier was not borrowed from the French hôtelier until 1905. Etymologists call such words doublets.

2. Note that the French hôtel has a diacritical mark called a circumflex on the o. This frequently occurs before a t and signifies an s which has been lost from the Old French. Thus, French pâte is the English paste, côte is the equivalent of English coast, and the French bête meaning "stupid" is related to the English word beast
3. Date "hotel" was first used: 1644
4. Dream Interpretation

   To dream of living in a hotel, denotes ease and profit.

   To visit women in a hotel, your life will be rather on a dissolute order.

   To dream of seeing a fine hotel, indicates wealth and travel.

If you dream that you are the proprietor of a hotel, you will earn all the fortune you will ever possess.

To work in a hotel, you could find a more remunerative employment than what you have.

To dream of hunting a hotel, you will be baffled in your search for wealth and happiness.

5. Hotels in fiction

Hotels have often been chosen by authors as the setting of their literary works, e.g. The Hotel New Hampshire. It is especially true of crime fiction (Agatha Christie's Evil Under the Sun, A Caribbean Mystery, At Bertram's Hotel; Cyril Hare's Suicide Excepted) and farces. Hotels also feature prominently in films (Grand Hotel, Room Service, Plaza Suite), television series, and songs, e.g. Hotel California.

6. Hotel is also the letter H in the NATO phonetic alphabet
7. Hotel was also the name of an American television program that aired on ABC from 1983 until 1988.
8. Synonyms within Context: Hotel (source: adapted from Roget's Thesaurus).

Abode House, mansion, place, villa, cottage, box, lodge, hermitage, rus in urbe, folly, rotunda, tower, chateau, castle, pavilion, hotel, court, manor-house, capital messuage, hall, palace; kiosk, bungalow; casa, country seat, apartment house, flat house, frame house, shingle house, tenement house; temple.

Assembly room, meetinghouse, pump room, spa, watering place; inn; hostel, hostelry; hotel, tavern, caravansary, dak bungalow, khan, hospice; public house, pub, pot house, mug house; gin mill, gin palace; bar, bar room; barrel house, cabaret, chophouse; club, clubhouse; cookshop, dive, exchange; grill room, saloon, shebeen; coffee house, eating house; canteen, restaurant, buffet, cafe, estaminet, posada; almshouse, poorhouse, townhouse. Hotel des Invalides; sanatorium, spa, pump room, well; hospice; Red Cross.

9. Etymologies containing "hotel": hostel.

Non-English Usage: "Hotel" is also a word in the following languages with English translations in parentheses.

Afrikaan (hotel), Albanian (hostel, hotel, inn), Asturian (hotel), Cebuano (hotel), Croatian (hotel), Czech (hotel), Danish (hotel), Dutch (hotel), Faeroese (hotel), Flemish (hotel), French (hotel), Frisian (hotel), Galician (hotel), German (hotel), Hawaiian (hotel), Hungarian (hotel), Indonesian (hotel), Italian (Hotel), Luxembourgish (hotel), Macedonian (hotel), Malay (hotel), Polish (hotel), Portuguese (hotel, inn), Portuguese Brazilian (hotel), Romanian (hotel, road house), Romansch (hotel), Serbo-Croatian (hotel), Slovene (hotel), Spanish (guesthouse, hotel, palace), Tagalog (hotel).

10. Use in Literature: Hotel

"Les Miserables" Hugo, Victor "*This done, he left the hotel and began to walk in the city" or "*Portrait of the Artist as a Young Man" Joyce, James "*One day he had stood beside her looking into the hotel grounds"*

11. Expressions: Hotel

Expressions using "hotel": apartment hotel / bespeak a room at a hotel / book accommodation at the hotel / cheap hotel / check out of a hotel / family hotel / Hotel and Restaurant Union / hotel bill / hotel business / hotel chain / hotel clerk / hotel desk clerk / hotel detective / hotel guest / hotel keeper / hotel lobby / hotel management / hotel manager / hotel occupancy / hotel plan / hotel porter / hotel reservation / hotel room / hotel runner / hotel servant / hotel staff / hotel thief / in the hotel lobby / leaving hotel without paying / luxury hotel / motor hotel / mountain hotel / proprietor of the hotel / resort hotel / seaside hotel / stay at a hotel without paying / stay in a hotel / temperance hotel / to hotel / villainous hotel.

Hyphenated Usage

Beginning with "hotel": hotel-based, hotel-building, hotel-casino, hotel-cum-clubhouse, hotel-cum-restaurant, Hotel-de-ville, Hotel-Dieu, hotel-guests, hotel-keeper, hotel-keepers, hotel-keeping, hotel-like, hotel-occupancy, hotel-owner, hotel-restaurant, hotel-room, hotel-schools, hotel-size, hotel-standard, hotel-style, hotel-type.

12. Ending with "hotel": casino-hotel, pub-turned-hotel, soundcheck-hotel-gig-hotel-travel-hotel, station-hotel, well-hotel.
13. Hotels in music

Hotel California of the band "The Eagles" touched on many themes, including innocence (and the loss thereof), addiction in general (and to drugs), death, the dangers, temptation and transient nature of fame, shallow relationships, divorce and loss of love, the end results of manifest destiny, and the "American Dream."

Members of Eagles have described the album as a metaphor for the perceived decline of America into materialism and decadence. In an interview with Dutch magazine

ZigZag shortly before the album's release, Don Henley said: “ This is a concept album, there's no way to hide it, but it's not set in the old West, the cowboy thing, you know. It's more urban this time (. . .) It's our bicentennial year, you know, the country is 200 years old, so we figured since we are the Eagles and the Eagle is our national symbol, that we were obliged to make some kind of a little bicentennial statement using California as a microcosm of the whole United States, or the whole world, if you will, and to try to wake people up and say 'We've been okay so far, for 200 years, but we're gonna have to change if we're gonna continue to be around.' ”

The album's final track, the epic "The Last Resort", was about the demise of society. Glenn Frey on the Hotel California episode of In the Studio with Redbeard explained about the track: “It was the first time that Don took it upon himself to write an epic story and we were already starting to worry about the environment...we're constantly screwing up paradise and that was the point of the song and that at some point there is going to be no more new frontiers. I mean we're putting junk, er, garbage into space now.”

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GLOBALIZATION AND EDUCATION...THEN AND NOW

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***Abstract**: Globalization has been one of the most debated topics in the world over the past few years. The process started approximately 1000 years ago. Rapid growth and poverty reduction in China, India, and other countries that were poor 20 years ago, has been one of the positive aspects of globalization. But globalization has also generated significant international opposition over concerns that it has increased inequality and environmental degradation.*
The paper presents a short history of globalization, its challenges and its effect on students and of course on the educational system. The paper discusses also the fact that quite often, globalization is seen as an exclusively Western phenomenon, which is understood, in most of the cases, as an aggressive force that often endangers indigenous cultures and ways of life.

***Keywords:** globalization, education, skills, youth, phenomenon*

The question – what is the most serious threat to our contemporary wellbeing, – might evoke the answer: “globalization.” Globalization, many of us believe, is a powerful juggernaut of transnational forces intent on devouring the world for profit, and destroying local cultures and environments in the process. However, from the perspective of history, things look different.

First, although globalization is often viewed as the consequence of new technologies and changing political priorities after the Cold War, it is in fact a much older force for change. A historical view of globalization, stressing the common heritage of humanity, enables us to contextualize changes such as industrialization without assigning superiority to the West. The result is a much more inclusive interpretation of globalization. Second, a historical interpretation enables us to treat globalization as a process facilitating the most radical transformation of societies that humans have ever experienced: a transformation called democratization. Globalization, then, is not about rampant capitalism, technology or homogenization. It is about the changed environments people create and manipulate as their societies globally interconnect environments that have become increasingly commercialized, urbanized, and democratized. [1]

Humans first experienced this transformative effect 1000 years ago, when the most advanced society of the time fueled continental interconnectedness. China’s trade surpluses fed a network of regional linkages that stretched across the world’s most populous continent.

But the momentum did not last. China succumbed to Mongolian warlords and much of Afro-Eurasia fell victim to the plague they carried across the continent. Like HIV/AIDS and SARS today, the plague was an early indicator that interconnectedness possesses real dangers. In the 14th century, China lost one-quarter of its population and Europe, one-third.

The resulting disruption encouraged European states to connect directly with China. In the process, they discovered the Americas, using its wealth to buy their way into the intra-Asian trade and establish a new Atlantic economy. Thus, human interconnectedness became globalized, and a whole new environment for human activities emerged. [2]

Today, the history of this early transformation is usually read in terms of civilizations and economic activity: Because Europeans initiated global networks, many observers stress European exceptionalism as its cause; because of the tremendous growth in commercial activities, many also give centrality to capitalism. But the transformation was much more extensive and destabilizing than these interpretations suggest. It accelerated the global distribution of plants and animals, transformed human diets, spawned rapid population growth, and stressed environments as land-use patterns changed and urbanization increased.

Authorities everywhere struggled to accommodate these unprecedented changes. Spain's rulers tried to convert their newfound wealth into the basis for hegemony within Europe, but in the end, power flowed to societies that gave space to wealth-generating merchants. This democratizing consequence of globalization challenged the interests of elites. They sought stability through exclusions of religion and race, as well as imperial and commercial monopolies. Frequently, they turned to war and conquest – outcomes that reduced global interconnectivity. Most elites little understood the transformations they confronted.

Industrialization is a case in point, normally portrayed as an example of British exceptionalism, or of Europe's "enlightenment." Rarely is it presented as an outcome of global production and trade in cotton and the expanding consumer markets globalization occasioned. Many analysts give centrality to technology in stimulating change. Yet human interconnectedness alone enabled industrialization to resonate so rapidly and globally. However, there can be no doubting the impact of what became a second wave of globalization. Industrialization enabled environments to carry larger populations, which in turn generated new social and political dynamics. More than ever before, technology generated huge profits, which made it desirable as an economic activity. Certainly the military power it generated, attracted the attention of states.

Yet most leaders still failed to grasp that security and well-being came from social empowerment, not conquest. They feared democratization and tried to reduce its impact. They sought colonial successes as alternatives. With competition increasingly drawn in Darwinian terms, they were prepared to go to war in order to retain hegemonic status at home and abroad. They failed, and World War I cost their nations tremendously – and not only in lives: It cost them the confidence that once energized the second wave of globalization. Economic collapse quickly generated a brutal depression. The resulting inward-looking economic policies simply reinforced the drive to empire and conquest that had already exacted a high price. During the 1930s and 1940s, they provoked a second round of bloodletting. [3]

From World War II, a very different third wave of globalization emerged. The demise of many former ruling classes created the democratic space for political stability in industrialized nations and for international cooperation. It also enabled the dissolution of empires. Thus decolonization, too, was a product of globalization. But decolonization could not guarantee meaningful participation in the new global environment for the emergent Third World. Colonialism left its peoples poorly equipped, and development strategies gave little weight to the democratic imperative. Consequently, a democratic global divide emerged that still holds the potential to destabilize human interconnectedness. In addition, the third wave began with a new global ideological division, an unprecedented arms race, and a destructive Soviet-American rivalry. It was not an auspicious start. However, as the Cold War ended, corporate transnationalism assumed center stage.

When postwar prosperity faltered in the 1970s, corporations exploited fears of recession to deregulate domestic economies and transform global regulatory systems to their advantage. As capital became more transnational, it harnessed a new generation of technological change to fashion global production networks. But it was not the only global force to survive the Cold War.

Postwar democratization had sent shockwaves of empowerment through industrialized societies. They reached deep into societies to transform working and domestic lives, family and social relationships, gender and race relations. They created wealthier, more educated, and longer-living populations able to connect with industry in new and innovative ways. Civil societies represented the symbiosis between economic growth and democratization. They stressed the role of human agency in development and generated alternative global goals. Such transformations demonstrate the dynamic character of the third wave of globalization. Human interconnectivity has always expanded the environments in which humans operate. But, as we have seen, it also generates challenges. [4]

Three challenges stand out: first, the challenge of extending and deepening democratization globally. Increases in inequalities, exacerbated by war and debt, have lost the third wave of globalization much of its legitimacy. However, like the empires of old, the industrialized world cannot survive as a world unto itself. Human interconnectivity makes that impossible. Second, there exists the environmental challenge of addressing issues of sustainability globally. Just as democracy cannot survive in a sea of poverty, so it cannot survive in an environmentally damaged and disease-ridden world. The third challenge is multicultural. We need to adjust to the diversity that globalization presents. With human migration greater and more rapid than at any time in the past, all forms of exclusivity risk instability. They deny common citizenship and collective responsibility as tools for sustainability in increasingly multicultural societies.

All three challenges represent divides that could cripple globalization and its human dynamic. Only democratization broadens the scope for wealth generation and capacity-building, and creates the skills needed to manage increasingly complex societies. [5]

But, as most of the critics admitted it, globalization defines our era. It is "what happens when the movement of people, goods, or ideas among countries and regions accelerates". [6] In recent years, globalization has come into focus, generating considerable interest and controversy in the social sciences, humanities, and policy circles and among the informed public at large. From terrorism to the environment, HIV-AIDS to Severe Acute Respiratory Syndrome (SARS), free trade to protectionism, population growth to poverty and social justice, globalization seems deeply implicated in nearly all of the major issues of the new millennium.

While globalization has created a great deal of debate in economic, policy, and grassroots circles, many implications and applications of the phenomenon remain virtual terra incognita. Education is at the center of this uncharted continent. We have barely started to consider how these accelerating transnational dynamics are affecting education, particularly precollege students' education. Instead, educational systems world-wide continue mimicking and often mechanically copying from each other and borrowing curricula (from trivial facts about history in middle school to trigonometry in high school), teaching methods ("chalk and talk"), and assessment tests (short answer and regurgitation). These practices would have been familiar to our forebears going to school two generations ago. [7]

Yet youth in school today, whether in Bali, Beijing, Beirut, Berlin, Boston, or Buenos Aires, will encounter a vastly different world from that of our grandparents. Throughout most human prehistory and history, the vectors that organized and gave meaning to human lives and human imaginaries were structured primarily by local geography and topology, local kinship and social organization, local worldviews and religions. Even a few hundred years ago, a minute in human evolutionary time, the lives of our ancestors were largely shaped by local economies, local social relations, and local knowledge. Prior to the transoceanic explorations and conquests, villagers were likely to be born, raised, and schooled (however shortly), to work, marry, reproduce, and be buried in the same locale. They were largely oblivious to changes taking place even a few hundred miles away. Then "the village was practically the beginning and end of his or her world: visitors were rare, few travelers passed by, and excursions from the village would, in all likelihood, have only been to the nearest market town. Contact with the outside world would have been the exception rather than the rule". [8]

Today the world tries to be another place. While human lives continue to be lived in local realities, these realities are increasingly being challenged and integrated into larger global networks of relationships. The forces of globalization are taxing youth, families, and education systems worldwide. All social systems are predicated on the need to impart values, morals, skills, and competencies to the next generation. [9] It seems obvious that the lives and experiences of youth growing up today will be linked to economic realities, social processes, technological and media innovations, and cultural flows. These global transformations, we believe, will require youth to develop new skills that are far ahead of what most educational systems can now deliver. New and broader global visions are needed to prepare children and youth to be informed, engaged, and critical citizens in the new millennium. A number of books has been developed around the idea that education will need both rethinking and restructuring if schooling is to best prepare the children and youth of the world to engage globalization's new challenges, opportunities, and costs.

Education's challenge will be to shape the cognitive skills, interpersonal sensibilities, and cultural sophistication of children and youth whose lives will be both engaged in local contexts, and responsive to larger transnational processes. We claim that several domains in particular will present the greatest challenges to schooling worldwide; among them, the domain of *complexity*.

Globalization engenders *complexity*. Throughout the world it is generating more intricate demographic profiles, economic realities, political processes, technology and media, cultural facts and artifacts, and identities. [10] Many countries are indeed undergoing intense demographic transformations. Sweden, a country of nine million people today, has a million immigrants, roughly half of them from the Muslim world. Economies likewise must adapt to the new, complex forces brought about by global capital. Local politics, too, are stretched in new ways – for example, when "absentee citizens" in the Diaspora exercise political power in the communities they left behind.

Globalization's increasing complexity necessitates a new paradigm for learning and teaching. The mastery and mechanical regurgitation of rules and facts should give way to a paradigm in which cognitive flexibility and agility win the day. The skills needed for analyzing and mobilizing to solve problems from multiple perspectives will require individuals who are cognitively flexible, culturally sophisticated, and able to work collaboratively in groups made up of diverse individuals.

In his contribution to this volume Howard Gardner claims that the complexity behind many of globalization's "big problems" requires deep disciplinary grounding *as well as* the ability to achieve multidisciplinary understandings, collaborations, and solutions. "Trends in our increasingly globalized society," writes Gardner, "have brought interdisciplinary concerns to the fore. Issues like poverty reduction, anti-terrorism, privacy, prevention of disease, energy conservation, ecological balance – and the list could be expanded at will – all require input from and syntheses of various forms of disciplinary knowledge and methods. Educational institutions seek, in their ways, to respond to the demand for this kind of skill; and the more adventurous students are attracted to studies that call for a blend of disciplinary expertise." Multitasking, learning how to learn, learning from failures, lifelong learning, and the ability to master and move across domains now have a premium. [11]

An education for globalization should therefore nurture the higher-order cognitive and interpersonal skills required for problem finding, problem solving, articulating arguments, and deploying verifiable facts or artifacts to substantiate claims. These skills should be required of children and youth who will, as adults, fully engage the larger world and master its greatest challenges, transforming it for the betterment of humanity – regardless of national origin or cultural upbringing.

Different authors have examined in their works how globalization has shaped the lives of the children of the world in and out of schools. The aim is to stimulate new thinking, research, and policy work in a domain that remains largely ignored by scholars of education. Millions of children and youth are growing up in a world where global processes are placing new demands on educational systems that are traditionally averse to change. There is virtually no scholarship on globalization and precollegiate education. [12] While there is some research on policy, administration, and curriculum that address globalization and primary and secondary education worldwide, generally these works fail to foreground how globalization is impacting the experiences of youth in and out of schools. Likewise, there is a small but growing literature on youth and globalization. Alas, most of these works fail to emphasize the role of education and schools in the lives of youth.

A number of researchers have begun to systematically examine how globalization is changing the lives of youth in Latin America and the Caribbean, in Arab countries, in sub-Saharan Africa and in Southeast Asia. More and more young people in these areas have access to global information; they copy the styles of U. S. teenagers (who themselves, as Jenkins informs us, borrow from youth elsewhere [13]), sing English-language songs, have more leisure opportunities for dating, and are more likely to be playing similar computer games. In many of these places, rural households shrink as a result of youth migrating to urban areas in search of work and other opportunities.

Gender roles are also transformed. While many observers see globalization as positive, promoting economic development and intercultural exchanges, there are also corrosive developments, such as globalization's threats to century-long traditions, religious identities, authority structures, values, and worldviews. It is increasingly obvious that in many corners of the world the winds of anti-globalization are blowing strong. [14]

Too often, globalization is seen as an exclusively Western phenomenon, an aggressive force that often endangers indigenous cultures and ways of life. But, as Robbie Robertson writes, this view is not simply reductive – it is inaccurate. "Globalization is not about rampant capitalism, technology, or homogenization," he writes, "It is about the changed environments people create and manipulate as their societies globally interconnect." [15] Robertson offers a broader historical perspective, tracing transformations from Chinese dynastic trade, through the bubonic plague, and up to today's telecommunications explosion. While interconnectivity has always expanded the environments in which humans operate, it has also generated substantial challenges. The most pressing issue, Robertson suggests, is in the necessity for stable, fair governance: "Only democratization broadens the scope for wealth generation and capacity-building, and creates the skills needed to manage increasingly complex societies."[16]

Notes

[1] *The Three Waves of Globalization: A History of a Developing Global Consciousness*, London & New York: Zed Books, 2003.
[2] Ibidem.
[3] Ibidem.
[4] Ibidem.
[5] Ibidem.
[6] Marcelo M. Suárez-Orozco, Desirée Baolian Qin-Hilliard, *Globalization: Culture and Education in the New Millennium*, University of California Press, Berkeley, CA, 2004. Page 1.
[7] Ibidem.
[8] Ibidem.
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OLD VS MODERN METHODS OF TEACHING ENGLISH

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Abstract: *The article focuses on different methods of teaching English and presents a parallelism between old teaching methods and new methods in teaching English. Teaching “traditional grammar” is one of the emphasized aspects. We also refer to the new teaching methods and their advantages and disadvantages for some of the learners and how well memorizing is entwined with what the new methods involve: learning and teaching using games, employing one language only (English) in the teaching process, or learners guessing the topic and the new grammar rule they are to study.*

Keywords*: teaching methods, traditional, perspective, approach, knowledge.*

Teaching a foreign language has always been fun to some, a challenge, a drag to others, passion, a crush, and to some maybe a pain. Surely, we are referring to teachers as well as students who, in turn, see teaching as interesting, mysterious, maybe boring. Some teachers prefer teaching the old way, some approach new methods. Both ways are good; still, each one has its own advantages and disadvantages. Still, what kind of teaching do students embrace? Do they like the old methods or the new ones?

When some of us were high-school students, foreign language teachers taught us English the old way. They didn’t resort to games or role-plays. They didn’t make us perform actions in order to memorize what was taught. Their only “ally” was the board, the chalk, the books, and last but not least, their gift and knowledge. Back at that time, we practiced a tense pattern by repeating it over and over until we fully understood and memorized it. This was what we call “traditional grammar”. “Traditional grammar" attempts, usually within a single language, to analyze and elucidate the constituents of any given well-formed sentence. The focus of attention is on surface structure, not meaning. The main benefit of "traditional grammar" is that it gives learners a basic understanding of the building blocks of language, which can help in improving their writing skills.

For effective teaching to take place a good method must be adopted by a teacher. A teacher has many options when choosing a style to teach by. The teacher may write lesson plans of their own, borrow plans from other teachers, or search online or within books for lesson plans. When deciding what teaching method to use, a teacher will need to consider students' background knowledge, environment, and learning goals. Teachers know that students learn in different ways but almost all of them will respond well to praise. Students have different ways of absorbing information and of demonstrating their knowledge. Teachers often use techniques which cater to multiple learning styles to help students retain information and strengthen understanding. A variety of strategies and methods are used to ensure that all students have equal opportunities to learn

Today, new teachers move toward new methods in their teaching. They practice teaching English by playing games with their students. The latter are encouraged to speak English even if they make errors.

Teachers teach them English by making them move around the classroom and look for different objects, asking them to make sentences in order to practice the new taught vocabulary or grammar.

By doing this, they practice the new vocabulary or grammar. This may be a good method when it comes to young learners. What about the adult learner?

Again, some may enjoy it, some may not. Does all this come down to their personalities? Perhaps. To some adult learners, learning English combined with playing it may be relaxing and enjoyable. To others, it may seem boring, a waste of time and knowledge. There are some aspects that we miss when applying this method. Some adult learners prefer learning new vocabulary or rules by listening and repeating. On the other hand, there are those who would vote in favor of writing something down. They feel that their visual memory can help them achieve their goal faster and better than learning only by listening. In this case, they need to put actions into words – that is noting down the rules that would later help them improve their speaking, listening, writing, or reading skills; it depends on which skill they aim or need to improve. Demonstrations are done to provide an opportunity in learning new exploration and visual learning tasks from a different perspective.

Another method would involve the "everything in English" method. We, as teachers, adore speaking the language we teach. We aim to make their speaking, listening, reading, and writing better by using only one language in the classroom-English. Yet, are we aware that some of our groups of learners may not fully understand our goal? If they don't understand what we say, how are they supposed to start speaking English and how are they supposed to achieve proficiency at the end? The "everything in English" method is appropriate for the advanced level students but it is definitely not good around beginners. Or maybe it is when it comes to very easy English.

Another issue would be making learners guess the new rule that the teacher is about to teach. We could call these jump-starts really interesting. They bring a little secrecy to the entire session; at least before the actual teaching session begins. The guessing "game" is meant to get the learners anticipate the topic and guess the new language aspect they are about to study. In this case, the teacher initiates a dialogue or can make a speech using the new grammar/vocabulary structures. The learners are supposed to spot the new structures and provide the topic for study.

Bearing all this in mind, we can conclude that methods, be they old or new, are effective within the teaching domain as long as learners attain a high level of mastering a foreign language. Most teachers today would embrace the new teaching methods: they are easier; they make learning and teaching relaxing. Yet, every teacher and learner should remember and appreciate any and every teaching/learning methods encountered and brought into play. The ultimate goal is to achieve the aimed objective.

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PERSON - RELATED AMERICAN SLANG

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Abstract: *Slang is the use of informal words and expressions to describe an object or condition.*
The definition of slang varies widely; however the generally accepted definition is of language which is very informal or much below the standard level of education - colloquially known as "street talk".

Keywords*: slang, culture, language, education*

Slang is the continual and ever-changing use and definition of words in informal conversation, often using references as a means of comparison or showing likeness.

Slang refers to vocabulary that is meant to be interpreted quickly but not necessarily literally. It is notable for its liveliness, humor, emphasis, brevity, novelty, and exaggeration. Most slang is faddish and ephemeral, but some words are retained for long periods and eventually become part of the standard language (e.g., phony, blizzard, and movie).

Slang can be born from any number of situations or ideas (the word slang itself has come to represent selling, especially of illegal drugs). We should keep in mind that slang comes from areas dominated by adolescents and young generations (such as music scene, school scene, drug scene, sex scene…)

The use of slang is frequently ridiculed by culturally-ignorant people who feel it is the product of insufficient education and believes it to be counter-evolutionary; of course, they couldn't be farther from the truth. Human language has been in a state of constant reinvention for centuries, and slang has been used and created by poets and writers of all sorts (even William Shakespeare has been credited for the upbringing of at least a couple of words).

Being exposed to movies in English, it is quite easy to spot words or expressions from this segment of the language, but it is not always easy to figure out their meaning. The present paper was intended to refer to some very used slang words related to persons, but because of the large number of words involved in such a list, I restrained the number of terms to those that involve types of behavior and personality.

1. **airhead**: stupid person.
   Sometimes he acts like an airhead!
2. **barf-out**: a displeasing person or affair.
   The person I met with yesterday was a barf-out.
3. **beat**: tired.
   At the end of the week we are all beat.

4. **boss**: excellent; great.
Your essay about globalization was boss!
5. **chicken**: coward.
When new ideas are brought into discussion, he is always a chicken.
6. **cool**: excellent; superb.
I have never met such a cool guy!
7. **couch potato**: a person who watches too much television.
She is so fat because she is a couch potato.
8. **deck**: to hit someone.
He was decked by the man because he said something about his girlfriend.
9. **dynamite**: powerful; excellent.
The night out was dynamite because of Jim's idea to go to the beach club.
10. **dirt**: extremely bad person.
His ex was dirt!
11. **fox**: attractive, alluring person.
Mickey was always a fox!
12. **geek**: someone who works too hard, is more intelligent than usual, and is slightly unattractive.
John was very proud to look like a geek.
13. **go bananas**: go slightly mad.
Because of Mary's rejection, Mike went bananas.
14. **gomer**: a dumb person.
Jim Carrey plays the part of a gomer in DUMB AND DUMBER.
15. **goof**: a silly and foolish person.
His appearance misguided us to think that he was a goof.
16. **goofy**: silly.
Don't be goofy!
17. **guts:** courage.
You don't have the guts to show he is wrong about it.
18. **hep**: sensible; informed.
She was always a hep student.
19. **hot** (1): popular.
The captain of the rugby team is hot among female students.
20. **hot** (2): sexy.
George Clooney is hot.
21. **hyper**: overly excited.
Finding out he won the first prize, he got Hyper.
22. **jerk**: stupid or annoying person.
Joanne's ex-boy friend was a real jerk!
23. **killer**: something exceptional or great.
The movie "The Fountain" is killer!
24. **knockout**: beautiful woman; handsome man.
Angelina Jolie is a knockout.
25. **loser**: a bungling and worthless person.
He is nothing but a loser.
26. **make waves**: cause problems.
I don't want to make waves but I will make myself heard!
27. **noid**: someone that's paranoid.

Don’t bring animal rights into discussion because he is noid.

28. **pissed (off)**: angry; upset.
She is definitely pissed off with you!
29. **pro**: someone who's good at something; professional.
No matter the problem, she is always acting like a pro.
30. **psycho**: crazy person.
He doesn’t act normally, he is definitely a psycho!
31. **rat:** a despicable person.
My ex- husband is a rat!
32. **riot**, a: something or someone very funny.
Charlie Chaplin is a riot!
33. **scum**: a despicable individual.
Why is he a scum? He sold his child!
34. **smarts**: intelligence.
You need smarts to write such a paper about ancient civilizations.
35. **smurfbrain**: a dumb or stupid person.
When caught in the act, he acted like a smurfbrain.
36. **sofa spud**: a person who watches too much television.
When European Championships begins, he transforms into a sofa spud.
37. **turkey**: dumb person.
Because she is the little one, she has always been treated like a turkey.
38. **uptight**: nervous; anxious.
Let him be; he is uptight because of the results in math.
39. **wimpy**: weak.
Don’t be a wimpy, fight them back!
40. **zero**: an unimportant person.
Why do you even bother to care: he is a zero!

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FIRST DAY ICEBREAKERS – GETTING TO KNOW YOUR STUDENTS

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Abstract: *Students often approach a new year and a new English teacher with some anxiety, some are not very talkative or relaxed at first, while others might even feel ill at ease when introducing and speaking about themselves. The purpose of this article is to present a set of helpful icebreakers which will put students at ease with the teacher and with each other. Icebreakers are short, helpful activities meant to make people more relaxed at the beginning of a course or fill those inevitable gaps during class. Icebreakers are the perfect first-day activities for an English class. Getting-to-know-you games and activities allow everyone in the class to get acquainted, help with team building and set a relaxed, fun tone to the first day.*

Key words: *icebreaker, method, interactive activities, speaking skills*

Interaction skills involve making decisions about communication, such as: what to say, how to say it, and whether to develop it, in accordance with one's intentions, while maintaining the desired relations with others.[1] For this reason, students might find introducing themselves and further talking about themselves quite a nerve racking task especially on the first day. Therefore, releasing some of this tension through interesting, fun, relaxing activities should make the object of every English teacher. The activities depicted here have as main purpose that of helping the teacher and the students break the ice during the first English class and get a good start to what can sometimes be a daunting task: speaking in English.

1. Learning student names

First of all, learning students' names is the basis of every good teacher – student relationship and an important aspect of class management. The present activity offers an idea for the beginning of school especially helpful for teachers who may have trouble learning new names. The teacher puts each student's name on a separate index card. Information such as phonetic pronunciation, gender, preferred nicknames is added. The teacher shuffles the cards after each round so that students can't anticipate their names. The teacher asks questions first, and then says the name so that all students listen to the question. The questions should be designed in such a way as to help the teacher get to know the new students, their likes, dislikes, family, experience etc.

This is also very helpful in assuring that all students are called on equally, not just the talkative ones. It also cuts down on the students who always want to be the ones to answer first.

[1] Martin Bygate, *Speaking*, Oxford University Press. 1993, p 6

Another fun and quick method of learning student names is to get them to say their name and with the first letter of their name to choose something they like doing to fit that. (e.g. My name is Robin and I like Reading. The next student starts with "Your name is Robin, you like Reading, My name is Dan I like Driving" this then builds up in a chain that goes around the class. By the end of the class the teacher will definitely know everyone's name.

2. What made you smile?

For breaking the incipient ice at the very beginning of a course or of a class, teachers could resort to a very easy but effective trick: smiling. It has been demonstrated that smiling not only relaxes the face muscles, but also helps relieve stress and tension. To further develop this strategy the teacher could use an ice-breaker that works very well with students at all levels. After entering the classroom with a very bright smile upon his/her face the teacher could ask "What made you smile today?" and let the students talk about the things that made their day/morning. To make the activity even more interesting the teacher could have the students work in pairs and speculate about what has made their colleague smile.

3. Mind mapping

Another activity meant for those students who are shyer to speak at the beginning appeals to the students' mind-mapping abilities. As necessary materials, there should be a blank piece of paper for each student as well as coloured markers. Each student needs to draw his/her name on the piece of paper - in as creative a way as possible, along with 5 things about themselves. It is very important to draw the items and not just write words. At the end of the allocated time they need to introduce themselves to those at their table using this as a springboard.

A variation to this activity could be asking the students to draw a humorous self portrait. They will add three things they like and three things they don't like on both sides of their portrait. When the task is finished, the teacher sticks each portrait on the classroom wall. The students can walk around admiring the portraits and guessing who is who. The teacher monitors the activity and can help start conversations by commenting upon the students' likes and dislikes, looking for coincidences or contrasts etc.

4. The lie detector

Telling and detecting lies can be an ideal way to start a course and get students to introduce themselves to the others[2]. Each person should write down, in a random order, 5 things about themselves that are true and 5 things that are false. Each person around the room reads their list in no particular order and the classmates have to guess what is true and what is false.

5. Things in common

Mingle activities such as this one are often designed to practice question asking and answering. In this case, the students must ask each other questions until they find

[2] *Recipes for Tired Teachers*, contributed by teachers associated with Pilgrims Language Courses, Addison-Wesley Publishing Co.Inc, 1985, p.45

three things that they have in common. They must be things that are not obvious. For example, they can't say “we both have black hair”. The teacher can exemplify by interviewing a student until he/she finds three things that they have in common. Students can repeat this several times and then report their findings to the class.

As Jill Hadfield once stated: “For a group to be harmonious and cohesive, it must have a definite sense of itself as a group, and the individuals who comprise it must have a sense of belonging to the group as well as a sense of their place within it.”[3] Therefore, as a variation to the previous activity, more advanced students could be asked to form groups according to their similarities. This works best with large groups. The purpose of this activity is to help students feel a sense of belonging. The activity can be further developed by having the students introduce themselves to that particular group they belong to and explain something about the association. It is important for the facilitator to watch for those who might not have a group. The teacher should either locate their group, or extend a group to include that person.

6. You're such a gossip

This is a very effective and enjoyable activity that can be used as a follow-up to a getting-to-know-you type of activity. As a warm-up the teacher could ask the students to work in pairs and say if they have ever heard any gossip about them. They could then be asked if they have ever invented any gossip about someone else and what was it. For the final activity students are divided in two groups. Each group of students will have to invent a gossip about the other group.

7. Who wrote what?

The following activity combines the writing and speaking skills and highly helps reduce stress in the sense that students are given time to write things about themselves and, therefore, are allowed to think for a while before speaking. This can be a very good way of getting to know students on the first day to school. To begin with, each student should have a blank sheet of paper. The teacher then asks one “getting-to-know-you” type of question for every student in the class. Students write the answer to the question and pass their paper to the next student. By the time this stage is over, each student should have a sheet of paper with a series of different answers, all written by different students. The teacher takes the papers in and distributes them at random. The students now have to mix and mingle in order to find out who wrote each answer on their sheet of paper, and write the name next to it. As a feedback the students could be asked to sum up what they know about each of their colleagues (e.g. “What do we know about Maria/Eric/Simon” etc

8. Getting to know the teacher

In order to achieve a good relationship with the students, the teacher should sometimes participate in the games and activities and let the students get to know him/her too. Students are usually very interested in the life of their teacher. A good way to share information is to bring a photo album into class. As the students look at the photos they each choose one and then think of three questions to ask the teacher about it. The teacher

[3] Jill Hadfield, *Classroom Dynamics*, Oxford University Press, 1994, p.72

gives them the answers and pairs them up. They then tell their partner about the photo and exchange photos. Next, they change partners and share the information they learned about the new photo. Partners can be changed several times. At the end of the session the teacher could ask questions about the photos because sometimes the information gets distorted to a highly humorous effect.

As a follow up, the teacher could get students to share information about themselves by imagining they have some representative pictures with them. Each new student is given a blank piece of paper which is folded in half. They have to imagine that they have brought four of their favorite photos from home, which may represent events, people or places that are important to them. Students will describe the imaginary pictures while the rest of the class has an opportunity to practice their questioning skills to find out more.

Another way to do this is to bring a variety of simple pictures in class. The pictures should be as simple as possible to add more personalized meanings to them (e.g. a baby, a car, an airplane, a toy, a TV, a ring). After each student has told the class a few things about himself/herself, he/she chooses a picture from an assortment of 15-20 pictures. The class has 3-5 guesses to try to figure out why the person chose the picture. If no one guesses correctly, then the student explains it.

9. How well do you "know" the teacher?

The teacher starts the first class by telling the students his/her name only. No other information about the teacher should be shared. Then the students are asked to write down 5 questions they would like to ask the teacher about whatever they may be interested in knowing about him/her. The teacher writes down each student's name on the board. Once they've written down their questions, one student is asked to come to the front of the class and play the role of the teacher, and will answer 5 of the students' questions, to the best of their guessing abilities. It is important to divulge the way the ice breaker works only after the students have finished writing their questions so as not to have any impact on the type of questions they will write down. Give a "point" to the student role-playing the teacher for each correct answer (or "close enough" answer) and an 'x' for an incorrect one. Each student takes turns guessing the answers to the students' questions until each student has answered 5 questions.

10. Student interviews

Instead of starting a class with all the students standing up and talking about themselves, which can be intimidating to some, the students could be asked to pair up and interview each other and introduce their colleague to the class. The variations can also be really fun. The students might interview each other on certain topics, which they will later explain to the class (e.g. the happiest day in their life, their favorite book or magazine, their favorite food, their plans for the future, a memorable vacation, etc.) With more advanced students the following variation can be very motivating. The idea of the exercise is for the students to mingle and introduce themselves to each other. The moment they have finished introducing themselves they assume the identity of the previous person. For example, Maria introduces herself to John. After that John is Maria and Maria is John. Next, the new John and Maria move on to other people. To add to this exercise the teacher can add simple questions for them to ask. The teacher offers in this way the conditions for a "normal every day verbal interaction."[4]

11. Find someone who

This is a good icebreaker for the first session, and also a very good way of reviewing grammar. Students are given slips of paper that contain several questions meant to help them find out things about their new colleagues. You can base your questions on students' level (e.g. Have you ever been to Europe? Have you ever eaten spaghetti? Have you ever met a VIP? etc.) They must talk to everyone in the class and find the appropriate person. Then, they must get as much related information as they can. Once they have finished, each student reads his/her question and identifies the person they chose. Follow-up questions can be asked afterwards.

An alternative version says that instead of giving the students the questions to ask, they could write a list of their own. There is a lot a teacher can learn about his/her students both from whom they want to find and how they answer.

The activities suggested here help students improve many of their skills. Firstly, after completing them, they will have learned a lot of information about the teacher and about their colleagues. Secondly, they will have asked a multitude of different question types, learning new question-expressions from each other as they go along. Thirdly, they will have been both asking and answering questions, the key to any conversation class. In addition, it makes them more comfortable, since with some activities they are not saying anything about themselves, which some students may be reluctant to do during the first class. Also, they are learning each other's names. This is greatly beneficial to the teacher as well, as it puts the students at ease and makes them feel better about expressing themselves in later classes.

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BRAINSTORMING ON THE PROCESS OF WRITING

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Abstract *If you remember the most important rule of writing then you will improve as a writer. Developing your own individual creative process and giving it time to work will make you a better writer. This is the case with most jobs -- whether you must write internal memos, correspond with clients, or help design sales materials. Writing beautiful prose and poetry is a talent. Writing effectively, however, is a skill that can be learned.*

Keywords: *writing, organization, grammatical devices, vocabulary, patterns.*

Why is it that for large numbers of English language students writing seems to pose great problems? Possibly for the same reason that large numbers of adult native speakers never achieve a high level of expressiveness in writing their first language. It is partly to do with the nature of writing itself. The writer is a lonely figure separated from the stimulus of the corrective learners. He is condemned to monologue; there is no one to help out, to fill the silences or to make encouraging noises.

Compared to speech, effective writing requires a number of things: a high degree of organization in the development of the ideas and information, a high degree of accuracy so that there is no ambiguity of meaning, the use of complex grammatical devices for focus and emphasis and a careful choice of vocabulary, grammatical patterns and sentence structures to create a style which is appropriate to the subject matter and the eventual readers.

Even those who are proficient writers in their first language have to acquire a wide language base from which to make choices. They may also find that confusing differences exist between the conventions of writing in their first language and English.

A good deal of writing in the English language classroom is undertaken as an aid to learning of new structures or vocabulary or to help students remember new items of language. In this context, the role of writing is little different from its role in any other subject; it allows students to see how they are progressing and to get feedback from the teacher, and it allows teachers to monitor and diagnose problems.

So there are many difficulties for the writers to struggle with when they ignore a simple rule. Once you embrace the fact that writing is a process rather than an event, once you recognize that the more time you give the process to work the better, then not only will writing be easier you will also write better.

Writing is a process. While that process varies somewhat based on the task and the individual writer, the basic steps it includes are the same no matter what. First is the initial brainstorming process. No actual writing takes place in this step although there may be some note taking or non-stop writing exercises. The more time you give yourself for this process then the easier the next step will be. Experiment with various forms of brainstorming and prewriting to determine which works best for you and your various writing tasks. What may work in one type of writing may not work as well with another. The more you experiment then the more likely you will find the optimum brainstorming process for you.

Second is the drafting process. That first rough draft should be a quick and painless draft. Your main goal at this point is simply to capture the fruits of your brainstorming in one document. Just write until you have tapped your brain. Do not hold yourself back by rewriting, revising, or editing. Do not pause to worry about spelling, grammar, punctuation, or word choice. If you are conscious that you will need to fill in gaps then simply hit return twice or write in all caps MORE LATER then move on. The important goal at this point is simply to capture your ideas in one place as quickly as possible. It does not have to be pretty and likely it will not be pretty, but it will be done.

Third is the revision process. This should take more than one draft to accomplish. Again, do not spend time worrying about spelling, grammar, punctuation, revising or editing. Fix the obvious errors that are distracting to you as you rework but that is not your main goal. Your main goal with this part of the process is to look at the big picture. Is your thesis clear and well supported? Are your ideas well organized and fully developed? Are there any gaps in the writing or logic? Do your ideas transition well from one to another?

Fourth is the editing process. Now is the time to worry about spelling, grammar, punctuation, and word choice. Zoom in your focus from the big picture to the sentence and paragraph level. This effort may take one or more drafts to polish your writing to the desired level.

If you are creating a more in-depth project then you may also need to add a step between brainstorming and drafting that includes research and organization which would make the writing process include five steps.

The most important part of creating your own individual writing process is to let it evolve as your skill grows. The more you refine and polish your process then the better the work you produce. The key to developing a successful writing process it to give yourself time - time to let your process evolve and time to let your writing develop. This means not to rush the development of your writing process. Let it evolve over many different projects. This also means not to rush your actual writing. Allow days to pass between various stages and drafts. The more time you allow passing then the more work your subconscious will do for you.

If you remember the most important rule of writing then you will improve as a writer. Developing your own individual creative process and giving it time to work will make you a better writer. This is the case with most jobs -- whether you must write internal memos, correspond with clients, or help design sales materials. Writing beautiful prose and poetry is a talent. Writing effectively, however, is a skill that can be learned.

Whether you are writing a memo to your co-worker or a report for your boss, you should decide what information you want to convey. We may always think of some tips to do it correctly.

Avoid wordiness. Say out loud what you are trying to write. Listen to how the words sound. Write for your audience. Use simple language. You don't want the reader to need a dictionary to decipher what you are trying to say. You should not try to impress your reader with your huge vocabulary. Chances are you will frustrate your reader instead. Most people are juggling several tasks at the same time, and are interested in receiving only necessary information. You are responsible for making this happen.

A cliché a day keeps the reader away - or at least it does not make him or her remember what you are saying. You want your writing to be memorable. Because we hear clichés often, we become desensitized to them. The words, then, are not uniquely associated with your writing. By no means pay attention to grammar. A good dictionary should be nearby, along with a thesaurus. A thesaurus will allow you to keep your writing fresh by helping you find a variety of words to use.

Proofreading is one of the most important things you can do. Since you probably do most of your writing on a computer, you have access to automated spelling and grammar checkers. Beware though - some words, used in the wrong context may be missed by computerized spell checkers. For example the sentence "To employees attended too meetings two learn about the gnu software," would pass through the spell check without any misspellings being detected. Have someone else proofread your document, if possible. If time allows, put your composition away, and proofread it later, or even better, the next day

Many people are intimidated by writing. Even so, there are times when writing is the best way to communicate, and often this is the only way to get your message across.

When writing, be mindful of the fact that once something is in written form, it cannot be taken back. Communicating in this way is more concrete than verbal communications, with less room for error and even less room for mistakes. This presents written communicators with new challenges, including spelling, grammar, punctuation, even writing style and actual wording.

Thankfully, today's technology makes memo, letter and proposal writing much easier by providing reliable tools that check and even correct misspelled words and incorrect grammar use. Unfortunately, these tools are not fail proof and will require your support, making your knowledge in this area important.

While the above tips cover the most common mistakes made when writing letters, memos and reports, they in no way cover everything you need to know to ensure your written communications are accurate and understood.

Perhaps the most important thing to remember when writing a letter is to check it thoroughly when it is completed. Even when you think it is exactly what you want, read it one more time. This "unwritten" rule holds true for everything you write – memos, letters, proposals, and so on.

Use both the grammar and spell check on your computer, paying very, very close attention to every word highlighted. Do not place total faith on your computer here. Instead, you should have both a printed dictionary and thesaurus nearby to double-check everything your computers editing tools highlight, as these tools are certainly not always reliable, for a variety of reasons.

When checking your written communications make sure the document is clear and concise. Is there anything in the written communication that could be misinterpreted? Does it raise unanswered questions or fail to make the point you need to get across?

Can you cut down on the number of words used? For instance, don’t use 20 words when you can use 10. While you do not want to be curt or abrupt, you do not want to waste the reader’s time with unnecessary words or phrases.

Is your written communication well organized? Does each idea proceed logically to the next? Make sure your written communications are easy to read and contain the necessary information, using facts where needed and avoiding information that is not relevant. Again, outline the course of action you expect, such as a return call or visit.

Close appropriately, making sure to include your contact information. While this may seem obvious, it is sometimes overlooked and can make your written communications look amateurish. This can diminish your chances of meeting your written communication’s goals.

The classroom needs to provide an environment in which students can experience being writers, thinking about purpose and audience, drafting a piece of writing, revising it, and sharing it with others.

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