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Jan Kučera

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SOLUTION IN LARGE OF CONTROL PROBLEM $\dot{x} = (A(1 - u) + Bu) x$

Jan Kučera, Praha

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Some notations used in this paper. The Euclidean *n*-dimensional space is denoted by E_n , its elements are written as matrices of *n* rows and one column. $e_i \in E_n$, i = 1, 2, ..., n, is that point which has the *i*-coordinate equal to one and the others equal to zero. We denote by δ_{ij} the Kronecker symbol equal to one if i = j and to zero if $i \neq j$. If $x \in E_n$, then we use the norm $||x|| = \sum_{i=1}^n |a_{ij}|$. which induces the norm of *n*-by-*n* matrix $A = (a_{ij})$ equal to $||A|| = \max_i \sum_{i=1}^n |a_{ij}|$.

The dimension of a (finite dimensional) vector space V one writes dim V. The symbol $\{x_1, x_2, ..., x_r\}$ denotes the linear hull of the elements $x_1, x_2, ..., x_r$ of a given linear space. If \Re is a linear space, then \Re \subset Ω means that \Re is a linear subspace of \Re .

The set of all elements $p \in P$, which have a property P(p), is denoted by $E(p \in P; P(p))$. The symbol O(t), $t \to 0$, represents a quantity, depending on t, which can be majorised by c|t|, where c > 0 is a sufficiently large constant, if t tends to zero.

If M is a given set, then \overline{M} is its adherence. Empty set one writes \emptyset . If φ is an one-to-one mapping, then φ^{-1} is an inverse mapping. We use the sings \cap , \cup to represent respectively the intersection and the union of sets. All measures and integrals, which occur later on, are meant in the Lebesgue sense.

The space of all *n*-by-*n* matrices one writes \mathfrak{E}_n . If $A, B \in \mathfrak{E}_n$, and $A_1, A_2, ..., A_r \in \mathfrak{E}_n$, we denote $[A, B] = BA - AB, [A_1, A_2, ..., A_r] = [A_1, [A_2, ..., [A_{r-1}, A_r] ...]]$.

We shall often meet the matrix $[A_1, A_2, ..., A_r]$, where $A_1 = A_2 = ... = A_{r-1}$. If there will be no danger to be mistaken then we shall write it briefly $[A_1, A_2, ..., A_r] = [A_1^{r-1}A_r]$.

The zero matrix is denoted by 0 and the unit matrix by E. A nonsingular matrix A possesses an inverse A^{-1} . The "bracket operation" [A, B] possesses the following properties: [A, B] + [B, A] = 0, $[A_1 + A_2, B] = [A_1, B] + [A_2, B]$, [A, B, C] + [B, C, A] + [C, A, B] = 0, [A, B, C, D] + [B, C, D, A] + [C, D, A, B] + [D, A, B, C] = [[A, C], [B, D]].

Definition 1. A connected set $S \subset E_n$ is called an r-dimensional manifold if for each $x \in S$ there exists an open non-empty set $G \subset E_r$ and an one-to-one mapping φ of G into S such that the following properties are satisfied:

- 1) $x \in \varphi(G)$,
- 2) $\varphi(G)$ is open in S,
- 3) the functional matrix $\partial \varphi / \partial \xi$ exists for all $\xi \in G$, is continuous and has rank r on G.

We often also say that the manifold S is given in an environ of the point x by the mapping φ .

If we have already an r-dimensional manifold $S \subset E_n$, then the set \overline{S} is called r-dimensional closed manifold. And finally, the set $S \subset E_n$ that contains only one element, one calls O-dimensional closed manifold.

Let us have an r-dimensional manifold S, given by a mapping φ , then every vector

$$\frac{\partial \varphi}{\partial \xi}\bigg|_{\xi=\varphi^{-1}(x)} \cdot \eta ,$$

where $\eta \in E_r$, $x \in S$, is called the tangent vector to the manifold S at the point x. The set of all tangent vectors to S at x is an r-dimensional vector space which we denote by T(x) and call tangent space to S at x.

It is obvious that the present definition of the dimension of a manifold and of a tangent space is independent on the choise of the mapping φ .

The set $S \subset E_n$ cannot be the adherence of two manifolds S_p , S_q with different dimensions p, q. If it would be so, then $\overline{S}_p = \overline{S}_q = S$. Let us suppose $S_p \cap S_q = \emptyset \Rightarrow S - S_p \supset S_q \Rightarrow \overline{S - S_p} \supset \overline{S}_q = S$. This is impossible as S_p is open in $S_p \subseteq S_p \cap S_q$, then there exist homeomorphisms φ, ψ of $G_p \subset E_p$, $G_q \subset E_q$ onto $\varphi(G_p) \subset S$, $\psi(G_q) \subset S$, respectively. Let $\varepsilon > 0$ be so small that the open sphere $K \subset E_n$, with centre at x and radius ε , has empty intersection with $(S - S_p) \cap (S - S_q)$ and further that $(K \cap S) \subset (\varphi(G_p) \cap \psi(G_q))$. Then the open set $\varphi^{-1}(K \cap S) \subset E_p$ is homeomorphic with the open set $\psi^{-1}(K \cap S) \subset E_q$ and p = q.

Definition 2. $\mathfrak{A} \subset \mathfrak{C}_n$. The mapping which assigns to every point $x \in E_n$ the vector space $\mathfrak{B}(x) = E(Ax; A \in \mathfrak{A})$ is called linear distribution created by \mathfrak{A} . As we will not investigate other distributions than linear, we will further omitt the adjective linear. Sometimes we write $\mathfrak{B}_{\mathfrak{A}}$ to underline that \mathfrak{B} has been created by \mathfrak{A} . If it holds: $A, B \in \mathfrak{A} \Rightarrow [A, B] \in \mathfrak{A}$, then \mathfrak{A} is called involutive. If for given distribution $\mathfrak{D}_{\mathfrak{A}}$ there exists an involutive space $\mathfrak{B} \subset \mathfrak{C}_n$ such that $\mathfrak{B}_{\mathfrak{A}} = \mathfrak{B}_{\mathfrak{B}}$, then $\mathfrak{B}_{\mathfrak{A}}$ is also called involutive.

Example. One distribution can be created by different matrix-spaces. Even there are non-involutive matrix-spaces which create an involutive distribution. If we have

two matrix-spaces $\mathfrak{A}_{1,2}$ which create the same distribution \mathfrak{B} , then the intersection $\mathfrak{A}_1 \cap \mathfrak{A}_2$ need not create \mathfrak{B} . To demonstrate it, let us put:

$$\mathfrak{A} = \left\{ \begin{pmatrix} 0 & 0 \\ 1 & 0 \end{pmatrix}, \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix}, \begin{pmatrix} 1 & 0 \\ 0 & -1 \end{pmatrix} \right\}; \quad \mathfrak{B} = \left\{ \begin{pmatrix} 1 & 0 \\ 0 & 1 \end{pmatrix}, \begin{pmatrix} 0 & -1 \\ 1 & 0 \end{pmatrix} \right\};$$

$$\mathfrak{C} = \left\{ \begin{pmatrix} 1 & 0 \\ 0 & 0 \end{pmatrix}, \begin{pmatrix} 0 & 0 \\ 0 & 1 \end{pmatrix}, \begin{pmatrix} 0 & 1 \\ -1 & 0 \end{pmatrix} \right\}; \quad \mathfrak{D} = \left\{ \begin{pmatrix} 1 & 0 \\ 0 & 0 \end{pmatrix}, \begin{pmatrix} 0 & 0 \\ 0 & 1 \end{pmatrix}, \begin{pmatrix} 0 & 1 \\ 1 & 0 \end{pmatrix} \right\}.$$
 Then
$$\mathfrak{A} \cap \mathfrak{B} = \left\{ \begin{pmatrix} 0 & -1 \\ 1 & 0 \end{pmatrix} \right\}.$$

 $\mathfrak{A}, \mathfrak{B}$ are involutive and $\mathfrak{C}, \mathfrak{D}$ are not. It holds $\mathfrak{B}_{\mathfrak{A}} = \mathfrak{B}_{\mathfrak{G}} = \mathfrak{B}_{\mathfrak{D}} + \mathfrak{B}_{\mathfrak{A} \cap \mathfrak{B}}$, while the spaces $\mathfrak{A}, \mathfrak{B}, \mathfrak{C}, \mathfrak{D}$ are different each other.

Example. The set $E(k; \exists \dim \mathfrak{B}(x) = k)$ may but need not contain all non-negative integers $k \le n$. Let us put

$$\mathbf{A} = \begin{bmatrix} 1 & 1 & 0 & 0 & \dots & 0 \\ 0 & 1 & 2 & 0 & \dots & 0 \\ 0 & 0 & 1 & 3 & \dots & 0 \\ \dots & \dots & \dots & \dots & \dots \\ 0 & 0 & 0 & 0 & \dots & (n-1) \\ 0 & 0 & 0 & 0 & \dots & 1 \end{bmatrix}, \qquad \mathbf{B} = \begin{bmatrix} 0 & 1 & 0 & \dots & 0 \\ 0 & 0 & 1 & \dots & 0 \\ \dots & \dots & \dots & \dots & \dots \\ 0 & 0 & 0 & \dots & 1 \\ 0 & 0 & 0 & \dots & 0 \end{bmatrix},$$

and write $B = B_1$, $[A, B_i] = B_{i+1}$, $i = 1, 2, ..., \mathfrak{A} = \{A, B_1, B_2, ..., B_{n-1}\}.$

Then $\mathfrak A$ is involutive and create a distribution $\mathfrak B$ for which holds: if $(x, e_i) \neq 0$, $(x, e_j) = 0$, j = i + 1, i + 2, ..., n, then $\mathfrak B(x) = \{e_1, e_2, ..., e_i\}$, i = 1, 2, ..., n. Here the symbol (x, e_i) means the scalar product.

1. SOLUTION OF EQUATION $\dot{x} \in \mathfrak{V}(x)$

Let $\mathfrak B$ be a given distribution in E_n and $\omega \in E_n$. In this paragraph we will find a necessary and sufficient condition for the existence of a solution of an equation

$$\dot{x} \in \mathfrak{V}(x), \quad x(0) = \omega,$$

and present the explicit form of that solution. What is to be understood as a solution of (1.1) is said in the following two definitions.

Definition 1.1. Each vector-function x, defined on an interval (0, T), $0 < T \le \infty$, local-absolutely continuous, which satisfies the conditions: 1) $x(0) = \omega$, 2) if there exists dx(t)/dt, then $dx(t)/dt \in \mathfrak{B}(x(t))$, is called a solution of (1.1).

Definition 1.2. Let \mathfrak{B} be a distribution and $S \subset E_n$ be a manifold. If for each $x \in S$ holds $T(x) = \mathfrak{B}(x)$, where T(x) is the tangent space to S at x, then S is called an integral manifold of the distribution \mathfrak{B} .

Lemma 1.1. Let $\mathfrak{A} \subset \subset \mathfrak{E}_n$ be involutive. Let us have a local-absolutely integrable matrix-function A(t), $t \geq 0$, and a matrix $B \in \mathfrak{A}$. Suppose that $A(t) \in \mathfrak{A}$ for all $t \geq 0$. Let us denote by X(t) that fundamental matrix of the equation $\dot{x} = A(t) x$, for which X(0) = E.

Then it holds $X^{-1}(t) BX(t) \in \mathfrak{A}, t \geq 0$.

We prove it in two steps: 1) Let $A \in \mathfrak{A}$ be a constant matrix, then $e^{-A}Be^A \in \mathfrak{A}$.

$$e^{-A}Be^{A} = \sum_{i,j=0}^{\infty} \frac{(-1)^{i}}{i! \, j!} A^{i}BA^{j} = \sum_{k=0}^{\infty} \frac{1}{k!} \sum_{i=0}^{k} (-1)^{i} \binom{k}{i} A^{i}BA^{k-i} =$$
$$= \sum_{k \ge 0} \frac{1}{k!} \left[\underbrace{A, A, ..., A, B}_{k} \right] \in \mathfrak{A}.$$

2) Let T>0. If A(t) is piecewise constant on $\langle 0,T\rangle$, then our statement is already proved by successive use of the first step. Let us choose a sequence of piecewise constant matrices $A_1(t), A_2(t), \ldots, t \in \langle 0,T\rangle$, which have all their values in $\mathfrak A$ and which tend asymptotically to A(t) on $\langle 0,T\rangle$. Let X_k be the fundamental matrix of the equation $\dot{x}=A_kx$, for which $X_k(0)=E, k=1,2,\ldots$ Then $X_k(t)-X(t)=\int_0^t A_k(\tau)(X_k(\tau)-X(\tau))\,\mathrm{d}\tau+\int_0^t (A_k(\tau)-A(\tau))\,X(\tau)\,\mathrm{d}\tau, \|X(t)\| \le \|E\|\exp\int_0^t \|A(t)\|\,\mathrm{d}\tau$. As X(t) is bounded on $\langle 0,T\rangle$, it holds $\lim_{k\to\infty}\int_0^t (A_k(\tau)-A(\tau))\,X(\tau)\,\mathrm{d}\tau=0$ uniformly on $\langle 0,T\rangle$. The asymptotic convergence of $\|A_k\|, k=1,2,\ldots$ implies $\sup_k\exp\int_0^t \|A_k\|\,\mathrm{d}\tau=K<\infty$. If we denote $b_k=\max_{t\in \langle 0,T\rangle}\|\int_0^t (A_k(\tau)-A(\tau))\,X(\tau)\,\mathrm{d}\tau\|$, we get the approximation $\|X_k(t)-X(t)\|\le b_k$. exp. $\int_0^t \|A_k(\tau)\|\,\mathrm{d}\tau\le K$. b_k , $t\in \langle 0,T\rangle$, where $\lim b_k=0$.

Now it holds $X_k^{-1}(t)$ B $X_k(t) \in \mathfrak{A}$, $k = 1, 2, ..., t \in (0, T)$, $\lim_{k \to \infty} X_k^{-1}(t)$ B $X_k(t) = X^{-1}(t)$ B $X(t) \in \mathfrak{A}$.

Lemma 1.2. Let $\mathfrak{A} \subset \mathfrak{E}_n$ be involutive and create the distribution \mathfrak{B} . Let $x \in \mathbb{E}_n$, $A_i \in \mathfrak{A}$, i = 1, 2, ..., k, $y = e^{A_1 t_1} e^{A_2 t_2} ... e^{A_k t_k} x$, where t_i , i = 1, 2, ..., k, are real numbers.

Then dim $\mathfrak{V}(y) = \dim \mathfrak{V}(x)$.

Proof. Let $Q \in \mathfrak{A}$ and $\mathfrak{B}(x) = \{P_1x, P_2x, ..., P_rx\}$, where $P_i \in \mathfrak{A}$, i = 1, 2, ..., r. For brevity we denote $\Phi = e^{A_1t_1}e^{A_2t_2}...e^{A_kt_k}$. It holds $Qy = Q\Phi x = \Phi(\Phi^{-1}Q\Phi)x$. According to lemma 1.1 $\Phi^{-1}Q\Phi \in \mathfrak{A}$ and $\Phi^{-1}Q\Phi x \in \{P_1x, P_2x, ..., P_rx\} \Rightarrow Qy \in \{\Phi P_1x, \Phi P_2x, ..., \Phi P_rx\}$. This implies dim $\mathfrak{B}(y) \leq \dim \mathfrak{B}(x)$.

If we change the signs of all numbers t_i ; i = 1, 2, ..., k, we get the inverse inequality.

Lemma 1.3. Let $\mathfrak{A} \subset \subset \mathfrak{E}_n$ be involutive and create the distribution \mathfrak{B} . Let $x \in E_n$, dim $\mathfrak{B}(x) = r$.

Then x is contained in an r-dimensional integral manifold S of the distribution \mathfrak{B} .

Moreover, if $\mathfrak{B}(x) = \{P_1x, P_2x, ..., P_rx\}$, where $P_i \in \mathfrak{A}$, i = 1, 2, ..., r, then S can be given in an environ of x by the mapping

(1.2)
$$\varphi(t_1, t_2, ..., t_r) = e^{P_1 t_1} e^{P_2 t_2} ... e^{P_r t_r} x, \quad t \in G,$$

where $G \subset E_r$ is an open set which contains the origin.

Proof. If dim $\mathfrak{B}(x) = 0$, then the statement is trivial, so let be dim $\mathfrak{B}(x) = r > 0$. Obviously $x = \varphi(0)$. Let us show that there is $\varepsilon > 0$ such that the mapping φ is one-to-one on the set $G_{\varepsilon} = E(t \in E_r; ||t|| < \varepsilon)$.

If it is not so, there would exist two sequences $t^1, t^2, ..., \tau^1, \tau^2, ...$, such that $t^k \to 0, \tau^k \to 0, t^k \neq \tau^k, \ \varphi(t^k) = \varphi(\tau^k), \ k = 1, 2, ...$ Let us put $t^k - \tau^k = \Delta^k, \ k = 1, 2, ...$, then it holds $0 = \varphi(t^k) - \varphi(\tau^k) = (P_1 \Delta_1^k + P_2 \Delta_2^k + ... + P_r \Delta_r^k) x + O(\|\Delta^k\|) \cdot O(\|t^k\| + \|\tau^k\|),$

$$\lim_{k \to \infty} \frac{1}{\|\Delta^k\|} \| (P_1 \Delta_1^k + P_2 \Delta_2^k + \ldots + P_r \Delta_r^k) x \| = 0.$$

The vectors $P_i x$, i = 1, 2, ..., r, are linearly independent so that it holds

$$\min_{\Delta \neq 0} \frac{1}{\|\Delta\|} \left\| \left(P_1 \Delta_1 + P_2 \Delta_2 + \dots P_r \Delta_r \right) x \right\| > 0.$$

This contrary proves the univalency of φ on G_{ε} .

The functional matrix $\partial \varphi / \partial t$ is continuous on E_n and at t = 0 has the rank equal to r. So we can choose $\varepsilon > 0$ so small that the matrix $\partial \varphi / \partial t$ has its rank equal to r for all $t \in G_{\varepsilon}$.

It remains to show that $\varphi(G_{\varepsilon})$ is an integral manifold of \mathfrak{B} . Obviously $P_i \varphi(t) \in \mathfrak{B}(\varphi(t))$, $i = 1, 2, ..., r, t \in E_r$. According to lemma 1.1

$$\frac{\partial \varphi}{\partial t_{i}} = e^{\mathbf{P}_{1}t_{1}} e^{\mathbf{P}_{2}t_{2}} \dots e^{\mathbf{P}_{i-1}t_{i-1}} \mathbf{P}_{i} e^{\mathbf{P}_{i}t_{i}} \dots e^{\mathbf{P}_{r}t_{r}} x =$$

$$= e^{\mathbf{P}_{1}t_{1}} \dots e^{\mathbf{P}_{i-1}t_{i-1}} \mathbf{P}_{i} e^{-\mathbf{P}_{i-1}t_{i-1}} \dots e^{-\mathbf{P}_{1}t_{1}} \varphi(t) \in \mathfrak{B}(\varphi(t)) ; \quad i = 1, 2, ..., r, \ t \in \mathbf{E}_{r}.$$

So we have got $\{\partial \varphi | \partial t_1, \partial \varphi | \partial t_2, ..., \partial \varphi | \partial t_r\} \subset \mathfrak{B}(\varphi(t))$, $t \in E_r$. For $t \in G_\epsilon$ there must hold the equality $\{\partial \varphi | \partial t_1, \partial \varphi | \partial t_2, ..., \partial \varphi | \partial t_r\} = \mathfrak{B}(\varphi(t))$, because we have chosen the set G_ϵ so that on G_ϵ the rank of the matrix $\partial \varphi | \partial t$ is equal to r and according to lemma 1.2 dim $\mathfrak{B}(\varphi(t)) = r$ for $t \in E_r$. This proves the equality.

Lemma 1.4. Let \mathfrak{B} be an involutive distribution. Let the point $x \in E_n$ is contained in two integral manifolds $S_{1,2}$ of \mathfrak{B} .

Then there exists an integral manifold S of \mathfrak{V} which is contained in the intersection $S_1 \cap S_2$ and contains the point x.

Proof. If dim $\mathfrak{B}(x) = 0$, then the statement is trivial. So we assume dim $\mathfrak{B}(x) = r > 0$. Let \mathfrak{B} be created by an involutive space $\mathfrak{A} \subset \mathfrak{S}_n$; $\mathfrak{B}(x) = \{P_1x, P_2x, ..., ..., P_rx\}$, $P_i \in \mathfrak{A}$, i = 1, 2, ..., r. Let the manifold S_3 be given by the formula (1.2). It will do to show that S_1 and S_3 have a common r-dimensional submanifold which contains the point x.

Let us choose $\Delta > 0$ so small that if $\|y - x\| < \Delta$, then the vectors $P_1 y, P_2 y, \ldots$, ..., $P_r y$ are linearly independent. Let the neighbourhood of x in S_1 be given by a mapping $\psi(\tau)$, $\|\tau\| < \delta$, $\psi(0) = x$. Let $\delta > 0$ is already chosen so small that the inequality $\|\tau\| < \delta$ implies $\|\psi(\tau) - x\| < \Delta$.

If $S_{1,3}$ have not a common submanifold, then there exists a sequence τ_1, τ_2, \ldots , such that $\|\tau_k\| < \delta$, $\psi(\tau_k) \notin S_3$, $k = 1, 2, \ldots, \tau_k \to 0$. The sequence τ_1, τ_2, \ldots has a partial sequence (let us denote it again by τ_1, τ_2, \ldots) such that there exists $\lim_{k \to \infty} \tau_k/\|\tau_k\| = \Omega$. Then there exists a differentiable rectificable curve $\tau = \tau(\vartheta)$, $\vartheta \in \langle 0, \vartheta_0 \rangle$, such that $\tau(0) = 0$, $\|\tau(\vartheta)\| < \delta$ for $\vartheta \in \langle 0, \vartheta_0 \rangle$, $\mathrm{d}\tau/\mathrm{d}\vartheta|_{\nu=0} = \Omega$, and such that there is a sequence $\vartheta_0 > \vartheta_1 > \vartheta_2 > \ldots > 0$ which satisfies $\tau(\vartheta_k) = \tau_k, k = 1, 2, \ldots$ For the curve $\Gamma_1(\vartheta) = \psi(\tau(\vartheta)), \vartheta \in \langle 0, \vartheta_0 \rangle$, it holds:

(1.3)
$$\frac{\mathrm{d}\Gamma_1(\vartheta)}{\mathrm{d}\vartheta} = \frac{\partial \psi}{\partial \vartheta} \frac{\mathrm{d}\tau}{\mathrm{d}\vartheta} = \sum_{k=1}^r p_k(\vartheta) P_k \Gamma_1(\vartheta) \qquad \vartheta \in \langle 0, \vartheta_0 \rangle.$$

As the vectors $P_k\Gamma_1(\vartheta)$, k=1,2,...,r, $\vartheta \in \langle 0,\vartheta_0 \rangle$, are continuous and linearly independent and as the vector-function $\partial \psi/\partial \tau$. $d\tau/d\vartheta$ is continuous on $\langle 0,\vartheta_0 \rangle$, too, the coefficients $p_k(\vartheta)$, k=1,2,...,r are also continuous functions on $\langle 0,\vartheta_0 \rangle$.

Now let us take a curve $\Gamma_2(\vartheta) = \varphi(t(\vartheta))$, $\vartheta \in \langle 0, \vartheta_0 \rangle$, where the function $t(\vartheta)$ has a continuous derivative on $\langle 0, \vartheta_0 \rangle$ and t(0) = 0, otherwise let it be in the meanwhile arbitrary.

$$\frac{\mathrm{d}\Gamma_{2}(\vartheta)}{\mathrm{d}\vartheta} = \frac{\partial\varphi}{\partial t} \cdot \frac{\mathrm{d}t}{\mathrm{d}\vartheta} = \sum_{s=1}^{r} \left(e^{\mathbf{P}_{1}t_{1}} \dots e^{\mathbf{P}_{s-1}t_{s-1}} \mathbf{P}_{s} e^{-\mathbf{P}_{s-1}t_{s-1}} \dots e^{-\mathbf{P}_{1}t_{1}}\right) \Gamma_{2}(\vartheta) \cdot \frac{\mathrm{d}t_{s}}{\mathrm{d}\vartheta} = \\
= \sum_{s=1}^{r} \left(\sum_{k=1}^{r} a_{ks}(t) \mathbf{P}_{k}\right) \Gamma_{2}(\vartheta) \frac{\mathrm{d}t_{s}}{\mathrm{d}\vartheta} = \sum_{k=1}^{r} \left(\sum_{s=1}^{r} a_{ks}(t) \frac{\mathrm{d}t_{s}}{\mathrm{d}\vartheta}\right) \mathbf{P}_{k} \Gamma_{2}(\vartheta).$$

The functions a_{ks} ; k, s = 1, 2, ..., r, are entire functions of the argument t and it holds $a_{ks}(0) = \delta_{ks}$; k, s = 1, 2, ..., r. If we now put

$$p_k(\vartheta) = \sum_{s=1}^{r} a_{ks}(t) \cdot \frac{dt_s}{d\vartheta}, \quad k = 1, 2, ..., r,$$

we get a system of equations from which we can calculate the derivatives $dt_s/d\vartheta$ s = 1, 2, ..., r, in an environ of the origin in E_r , and we get another system which

has obviously a unique solution satisfying the initial condition t(0) = 0. So there exists $\tilde{\vartheta} > 0$ such that on $\langle 0, \tilde{\vartheta} \rangle$ the vector-functions $\Gamma_{1,2}$ solve the linear system of equations (1.3) and satisfy the initial condition $\Gamma_1(0) = \Gamma_2(0) = x$. Hence the both functions are on $\langle 0, \tilde{\vartheta} \rangle$ identical, what is the sought contrary.

We have proved that there exists a neighbourhood $G \subset E_r$ of the origin in E_r such that $\psi(G) \subset S_3$. Let G be so small that the ranks of the matrices $\partial \psi/\partial \tau$, $\partial \varphi/\partial \tau$ are both equal to r for all $\tau \in G$. Then there exists a neighbourhood $G_0 \subset G$ of the origin in E_r such that the equation $\psi(\tau) - \varphi(t) = 0$ has the unique solution $\tau = \tau(t)$ for $t \in G_0$. And the set $\psi(\tau(G_0)) = \varphi(G_0)$ is the sought common r-dimensional submanifold of the manifolds S_1 , S_3 .

Lemma 1.5. Let \mathfrak{V} be a distribution in E_n and let each point $x \in E_n$ be contained in an integral manifold S_x of \mathfrak{V} . Then \mathfrak{V} is involutive.

Proof. Let us put $\mathfrak{U} = E(A \in \mathfrak{E}_n; Ax \in \mathfrak{D}(x) \text{ for all } x \in E_n)$, choose $x \in E_n$ and $A, B \in \mathfrak{U}$. The point x is contained in a manifold S_x . According to lemma 1.4 it holds $e^{-Bt}e^{-At}e^{Bt}e^{At}x \in S_x$ for sufficiently small real t. This implies

$$\lim_{t \to 0} t^{-2} (e^{-Bt} e^{-At} e^{Bt} e^{At} - E) x = [A, B] x \in T(x)$$

and $[A, B] x \in \mathfrak{V}(x)$. As it holds for each $x \in E_n$, it is $[A, B] \in \mathfrak{A}$.

Lemma 1.6. Let $\mathfrak{A} \subset \subset \mathfrak{E}_n$ be involutive and create the distribution \mathfrak{B} . Let for matrices $B, C \in \mathfrak{E}_n$ be $Bx \in \mathfrak{B}(x)$, $Cx \in \mathfrak{B}(x)$ for all $x \in E_n$.

Then
$$[B, C] x \in \mathfrak{B}(x)$$
 for all $x \in E_n$.

Proof follows immediately from the proof of lemma 1.5.

Supplement to definition 2. Let $\mathfrak B$ be an involutive distribution. Let us form the space

(1.4)
$$\mathfrak{A} = E(A \in \mathfrak{E}_n; \ Ax \in \mathfrak{B}(x) \text{ for all } x \in E_n)$$

Then according to lemma 1.6 the space $\mathfrak A$ is involutive. So by (1.4) we can uniquely to each involutive distribution $\mathfrak B$ determine the involutive space $\mathfrak A \subset \mathfrak F_n$ such that $\mathfrak B = \mathfrak B_{\mathfrak A}$.

The definition of being involutive for a given distribution might be now given as follows: A given distribution \mathfrak{B} is called involutive iff the space (1.4) is involutive.

Theorem 1.1. Let \mathfrak{B} be a distribution in E_n . Let us denote $Z_r = E(x \in E_n; \dim \mathfrak{B}(x) = r)$, r = 0, 1, ..., n, and take one connected component Z of a set Z_r .

Then each $x \in Z$ is contained in a unique r-dimensional integral manifold S_x , of the distribution \mathfrak{B} , which is maximal in the sense of inclusion of sets, if and only if \mathfrak{B} is involutive.

Moreover, if \mathfrak{B} is created by an involutive space $\mathfrak{A} \subset \subset \mathfrak{E}_n$, $\mathfrak{B}(x) = \{P_1x, P_2x, ..., ..., P_rx\}$, $P_i \in \mathfrak{A}$, i = 1, 2, ..., r, then the integral manifold S_x is given by the mapping (1.2).

Proof. 1) The necessity follows from lemma 1.5. 2) Let $\mathfrak B$ be involutive. Then according to lemma 1.3 each point $x \in Z$ is contained in an integral manifold of $\mathfrak B$. Let us choose a new topology in Z like in [2]. That topology consists of all sets which can be represented as unions of integral manifolds of $\mathfrak B$. Then the sought manifold S_x is that component of Z which contains the point x.

The supplement is an immediate consequence of lemma 1.3.

Theorem 1.2. Let \mathfrak{B} be an involutive distribution in E_n and $\omega \in E_n$. Then the integral manifold S_{ω} is the set of all points $x \in E_n$ for which there exists a solution x(t) of the equation (1.1) and T > 0 such that x = x(T).

Proof. 1) From the proof of lemma 1.4 it follows that all points lying on a solution of (1.1) are contained in S_{ω} .

2) Let $x \in S_{\omega}$, then x can be linked up with ω by a finite chain of integral manifolds S_i , i=1,2,...,k, given by the formula (1.2), $\omega \in S_1$, $S_i \cap S_{i+1} \neq \emptyset$, $i=1,2,...,\ldots,k-1$, $x \in S_k$. And obviously if we have two arbitrary points $x_{1,2} \in S_1$, then there exists a solution x(t) of the equation $\dot{x} \in \mathfrak{B}(x)$, $x(0) = x_1$, and a number T > 0 such that $x(T) = x_2$. This completes the proof.

2. CONTROL PROBLEM

In this paragraph we will investigate the equation

(2.1)
$$\frac{d}{dt} x(t) = (A(1 - u(t)) + B u(t)) x(t), \quad x(0) = \omega, \quad t \ge 0,$$

where A, $B \in \mathfrak{C}_n$, $\omega \in E_n$ and u is a measurable function on $(0, \infty)$, values of which lie in (0, 1) for all $t \ge 0$.

The matrices $A, B \in \mathfrak{E}_n$ will be fixed. As the matrix B - A will appear very frequently we shall consistently denote it by C.

The set of all functions measurable on $\langle 0, \infty \rangle$, values of which lie in $\langle a, b \rangle$, a < b, we denote by M(a, b). The function $u \in M(0, 1)$ one call the control. The solution of (2.1), which corresponds to a given control $u \in M(0, 1)$, we denote by x(t, u). And at last we denote by X(t) the fundamental matrix of (2.1) for which X(0) = E. Here we do not indicate explicitly the dependence of X(t) on the control u, because it will be still clear what control will be dealt with.

Definition 2.1. The smallest linear involutive space of n-by-n matrices, which contains the matrices A, B, we denote by \mathfrak{A} (or by $\mathfrak{A}(A, B)$) and the distribution created by \mathfrak{A} we denote by V.

Definition 2.2. By \mathfrak{B} (or by $\mathfrak{B}(A, B)$) we denote the smallest linear space, of n-by-n matrices, which contains the matrix C and with each $P \in \mathfrak{B}$ contains also both matrices [A, P] and [B, P].

The distribution created by $\mathfrak B$ we denote by $\mathscr V$.

Lemma 2.1. $\mathfrak{B}(A, B)$ is involutive.

Proof. Let us call the matrix $P \in \mathfrak{E}_n$ elementary of grade p, if there exists a sequence of matrices $P_1, P_2, ..., P_{p-1}$, where $P_i = A$ or $P_i = B$, i = 1, 2, ..., p-1 such that $P = [P_1, P_2, ..., P_{p-1}, C]$. We do not care that an elementary matrix could have different grades.

Evidently \mathfrak{B} is the linear hull of all elementary matrices. To prove the involutivity of \mathfrak{B} , it is sufficient to show that if there are given elementary matrices P, Q, with grades p, q, respectively, then the matrix [P, Q] is a linear combination of elementary matrices of grades less or equal to p + q.

If p + q = 2, then [P, Q] = 0, and the statement is obvious. Let us assume that our statement holds for all elementary matrices which have the grades p, q, where p + q < r. Now, let matrices P, Q have the grades p, q, respectively, and p + q = r. If p = 1, then obviously [P, Q] is elementary. If p > 1, then, without loss of generality, let be P = [A, R], where R is elementary of grade p - 1.

$$[P, Q] = [[A, R], Q] = [A, [R, Q]] + [R, [Q, A]].$$

The matrix [R,Q] is a linear combination of elementary matrices of the grades at p+q-1. As [A,Q] is elementary and R has its grade less then p we get by means of the mathematical induction with respect to p that also [R,[Q,A]] is a linear combination of elementary matrices of grades less or equal to p. This completes the induction with respect to p.

Lemma 2.2. A, $B \in \mathfrak{C}_n$; $\gamma_1, \gamma_2 \in E_1$; $\gamma_1 \neq \gamma_2$. Let $\mathfrak{M} \subset \subset \mathfrak{C}_n$ be the smallest linear space with the properties:

- 1) $C \in \mathfrak{M}$
- 2)' $P \in \mathfrak{M} \Rightarrow [A + \gamma_i C, P] \in \mathfrak{M}, i = 1, 2.$

Then $\mathfrak{M} = \mathfrak{B}(A, B)$.

Proof.
$$P \in \mathfrak{M} \Rightarrow [A, P] = (1/(\gamma_2 - \gamma_1)(\gamma_2[A + \gamma_1C, P] - \gamma_1[A + \gamma_2C, P]) \in \mathfrak{M},$$
 $[B, P] = (1/(\gamma_2 - \gamma_1)((\gamma_2 - 1)[A + \gamma_1C, P] - (\gamma_1 - 1)[A + \gamma_2C, P]) \in \mathfrak{M}.$

Lemma 2.3.
$$\mathfrak{A}(A, B) = \{A, \mathfrak{B}(A, B)\}, V(x) = \{Ax, \mathscr{V}(x)\} \text{ for all } x \in E_n$$
.

Proof follows immediately from the proof of lemma 2.1.

Definition 2.3. $\omega \in E_n$, $T \ge 0$. Then the set of all points $x \in E_n$ for which there exists such $u \in M(0, 1)$ that x = x(T, u) (x(t, u)) is a solution of (2.1), we denote by $\mathscr{S}_{\omega}(T)$ and we write $\bigcup_{t \in \langle 0, T \rangle} \mathscr{S}_{\omega}(t) = S_{\omega}(T)$.

Lemma 2.4. The sets $\mathscr{S}_{\omega}(T)$, $S_{\omega}(T)$ are compact and connected for all $\omega \in \mathbb{E}_n$, $T \geq 0$.

Proof. 1) $S_{\omega}(T)$ is bounded. Denote $\alpha = \max_{u \in \langle 0, 1 \rangle} \|A + uC\|$, then for every solution of (2.1) it holds $\|x(t, u)\| \le \|\omega\| + \alpha \int_0^t \|x(\tau, u)\| d\tau \Rightarrow \|x(t, u)\| \le \|\omega\| \cdot e^{\alpha t}$.

2) $S_{\omega}(T)$ is compact. Let a sequence $x_i \in S_{\omega}(T)$, $i=1,2,\ldots$ be given. Then there exist $t_i \in \langle 0,T\rangle$ and $u_i \in M(0,1)$, $i=1,2,\ldots$ such that $x_i=x(t_i,u_i)$, $i=1,2,\ldots$ There is a convergent subsequence of t_1,t_2,\ldots Let its limit be t_0 and let us suppose, without loss of generality, that already t_1,t_2,\ldots is convergent to t_0 . The sequence $x(t,u_i)$, $t \in \langle 0,t_0 \rangle$, is uniformly bounded due to boundeness of $S_{\omega}(T)$ and uniformly continuous. So there exists a subsequence (let it be again the original sequence) that converges uniformly to a function x(t), $t \in \langle 0,t_0 \rangle$.

The sequence $u_1, u_2 \ldots$ is bounded in the space $L_2(0, t_0)$ of all square-integrable functions on $\langle 0, t_0 \rangle$ and we can choose from it a subsequence (let it be again u_1, u_2, \ldots) weakly convergent (in $L_2(0, t_0)$) to a function $u \in L_2(0, t_0)$. Let us show that $u \in M(0, 1)$. If $\varepsilon > 0$, denote χ the characteristic function of the set $Q = E(t \in \langle 0, t_0 \rangle; u(t) > 1 + \varepsilon)$. Then $\mu(Q) \geq \int_0^{t_0} u_i(\tau) \chi(\tau) d\tau$, $i = 1, 2, \ldots, \mu(Q) \geq \int_0^{t_0} u(\tau) \chi(\tau) d\tau \geq (1 + \varepsilon) \mu(Q) \Rightarrow \mu(Q) = 0$. In the same way we can prove that $u(t) \geq 0$ almost everywhere on $\langle 0, t_0 \rangle$.

It holds: $\left|\int_0^{t_0} \left(u_i x_i - ux\right) \, d\tau\right| \le \left|\int_0^{t_0} u_i (x_i - x) \, d\tau\right| + \left|\int_0^{t_0} \left(u_i - u\right) x \, d\tau\right| \to 0$ $\int_0^{t_0} \left(A + u_i C\right) x_i \, d\tau \to \int_0^{t_0} \left(A + uC\right) x \, d\tau = x(t_0, u) - \omega$. We have got x(t) = x(t, u) for $t \in \langle 0, t_0 \rangle$, hence $x(t_0) = x(t_0, u) \in S_{\omega}(T)$. The proof of the compacteness of $\mathcal{S}_{\omega}(T)$ is similar.

3) $S_{\omega}(T)$ is obviously connected. We prove that also $\mathscr{S}_{\omega}(T)$ is connected.

Let x(T, u), $x(T, v) \in \mathscr{S}_{\omega}(T)$, then the set $E(x(T, u(1 - \lambda) + v\lambda); \lambda \in \langle 0, 1 \rangle)$ is connected, included in $\mathscr{S}_{\omega}(T)$ and includes both points x(T, u), x(T, v).

Let us fix $\delta \in (0, \frac{1}{2})$, $u \in M(\delta, 1 - \delta)$ and put in (2.1) instead of the control u the control $u + \varepsilon v$, where $v \in M(-1, 1)$, $\varepsilon \in (0, \delta)$. Obviously $u + \varepsilon v \in M(0, 1)$ and we get

(2.2)
$$\dot{x} = (A + Cu) x + \varepsilon Cvx, \quad x(0) = \omega.$$

The solution of (2.2) is an analytic function of the parameter ε .

$$(2.3) x(t, u + \varepsilon v) = x_0(t, v) + \varepsilon x_1(t, v) + \varepsilon^2 x_2(t, v) + \dots$$

(2.4)
$$\dot{x}_0 = (A + Cu) x_0, \quad x_0(0, v) = \omega,$$

(2.5)
$$\dot{x}_k = (A + Cu) x_k + Cv x_{k-1}, \quad x_k(0, v) = 0, \quad k = 1, 2, ...$$

Let us put $\alpha = \max_{u \in \langle 0, 1 \rangle} \|A + Cu\|$ and estimate $\|x_k(t, v)\|$, k = 1, 2, ...

$$||x_0(t,v)|| \le ||\omega|| \cdot e^{\alpha t}$$

$$\begin{split} \frac{\mathrm{d}}{\mathrm{d}t} \|x_k\| & \leq \alpha \|x_k\| + \|C\| \cdot \|x_{k-1}\| \Rightarrow \|x_k(t,v)\| \leq \|C\| \int_0^t e^{\alpha(t-\tau)} \|x_{k-1}\| \, \mathrm{d}\tau \\ \|x_k(t,v)\| & \leq \|\omega\| \cdot \|C\|^k \frac{t^k}{k!} e^{\alpha t} \,, \quad k = 0, 1, 2, \dots \\ \|x(t,u+\varepsilon v)\| & \leq \sum_{k \geq 0} \varepsilon^k \|x_k(t,v)\| \leq \|\omega\| \cdot e^{(\alpha+\varepsilon\|\varepsilon\|)t} \,. \end{split}$$

Thus the serie (2.3) is locally uniformly absolutely convergent. We can write

(2.6)
$$x_1(T, v) = X(T) \int_0^T X^{-1}(t) Cx_0(t, v) v(t) dt =$$

$$= X(T) \int_0^T X^{-1}(t) CX(t) v(t) dt \omega.$$

Lemma 2.5. The set $K_u(T)$, T > 0, of all vectors (2.6), where $v \in M(-1, 1)$, is convex and symetric with the centre at origin.

Proof follows immediately from (2.6).

Lemma 2.6. The linear hull of $K_u(T)$ is identical with the linear hull of the vectors

(2.7)
$$X(T) X^{-1}(t) CX(t) \omega, \quad t \in \langle 0, T \rangle.$$

Proof. By integration of (2.7) we cannot leave the linear hull. On the contrary let us fix $t \in (0, T)$ and put

$$v_{\alpha}(\tau) = \begin{cases} 1 \\ 0 \end{cases} \quad \text{for} \quad \begin{cases} \tau \in \langle t - \alpha, \ t + \alpha \rangle \cap \langle 0, T \rangle \\ \tau \in \langle 0, T \rangle - \langle t - \alpha, \ t + \alpha \rangle \end{cases}, \quad \alpha > 0$$

Then

$$\lim_{\alpha \to 0+} (2\alpha)^{-1} X(T) \int_{0}^{T} X^{-1}(\tau) CX(\tau) v_{\alpha}(\tau) d\tau \omega =$$

$$= \varphi(t) X(T) X^{-1}(t) CX(t) \omega,$$

where

$$\varphi(t) = \begin{cases} 1 \\ \frac{1}{2} \end{cases} \quad \text{for} \quad \begin{cases} t \neq 0, \ t \neq T \\ t = 0 \text{ or } t = T \end{cases}.$$

Lemma 2.7. Let T > 0, $u_{\lambda} \in M(0, 1)$ for all $\lambda \in \langle 0, \lambda_0 \rangle$, $\lambda_0 > 0$. If $u_{\lambda} \to u_0$ asymptotically on $\langle 0, T \rangle$, $\lambda \to 0$, then $X(t, u_{\lambda}) \to X(t, u_0)$ uniformly on $\langle 0, T \rangle$.

Here $X(t, u_{\lambda})$ is the fundamental matrix-function of equation (2.1), where u is replaced by u_{λ} , and $X(0, u_{\lambda}) = E$.

Proof is contained in the second part of the proof of lemma 1.2.

Lemma 2.8. T > 0, $\delta \in (0, \frac{1}{2})$, $u \in M(\delta, 1 - \delta)$. Let the function u be not constant on (0, T) (not equivalent with a constant function), then

(2.8)
$$\mathscr{V}(x(T,u)) \subset \bigcup_{r=1}^{\infty} r \cdot K_u(T).$$

Proof. According to lemma 2.6 the vectors (2.7) are contained in $\bigcup_{r=1}^{\infty} r \cdot K_u(T)$. For almost all $t \in \langle 0, T \rangle$ has the function $\int_0^t u(\tau) d\tau$ the derivative equal to u(t). Let $G \subset \langle 0, T \rangle$ be the set where it is not true. Let us take $t_1 \in \langle 0, T \rangle - G$, then we can write $X(T) X^{-1}(t) CX(t) \omega = X(T) X^{-1}(t_1) (X(t_1) X^{-1}(t) CX(t) X^{-1}(t_1)) X(t_1)$. $X^{-1}(T) x(T, u)$. Let us, for brevity, denote $X(t) X^{-1}(t_1) = Y(t)$ and divide the proof into two parts. In the first, resp. second, part we show that the linear hull L, resp. L_1 , of matrices $Y^{-1}(t) CY(t)$, resp. $X(T) X^{-1}(t) CX(t) X^{-1}(T)$, where t ranges through entire $\langle 0, T \rangle$, contains the space $\mathfrak{B}(A, B)$.

1) For $t = t_1$ we get $C \in L$. Let all elementary matrices from $\mathfrak{B}(A, B)$ of the grades equal at most to (k-1) (the definition of the elementary matrix and its grade is given in the proof of lemma 2.1) belong to L. The matrix-function Y satisfies the equation $\dot{Y} = (A + uC) Y$, $Y(t_1) = E$ and for $t \in (t_1, T)$ we can write

$$Y(t) = E + \int_{t_{1} < \tau_{1} < t} (A + u(\tau_{1}) C) d\tau_{1} + \dots +$$

$$+ \int_{t_{1} < \tau_{k} < \dots < \tau_{1} < t} (A + u(\tau_{1}) C) \dots (A + u(\tau_{k}) C) d\tau_{k} \dots d\tau_{1} +$$

$$+ \int_{t_{1} < \tau_{k+1} < \dots < \tau_{1} < t} (A + u(\tau_{1}) C) \dots (A + u(\tau_{k+1}) C) Y(\tau_{k+1}) d\tau_{k+1} \dots d\tau_{1}$$

$$\frac{d}{dt} Y^{-1}(t) = -Y^{-1}(t) \left(\frac{d}{dt} Y(t)\right) Y^{-1}(t) = -Y^{-1}(t) (A + u(t) C),$$

$$Y^{-1}(t_{1}) = E,$$

$$Y^{-1}(t) = E - \int_{t_{1} < \tau_{1} < t} (A + u(\tau_{1}) C) d\tau_{1} + \dots +$$

$$+ (-1)^{k} \int_{t_{1} < \tau_{k} < \dots < \tau_{1} < t} (A + u(\tau_{k}) C) \dots (A + u(\tau_{1}) C) d\tau_{k} \dots d\tau_{1} +$$

$$+ (-1)^{k+1} \int_{t_{1} < \tau_{k+1} < \dots < \tau_{1} < t} Y^{-1}(\tau_{k+1}) (A + u(\tau_{k+1}) C) \dots$$

$$\dots (A + u(\tau_{1}) C) d\tau_{k+1} \dots d\tau_{1}$$

$$Y^{-1}(t) CY(t) = C + \int_{t_{1} < \tau_{1} < t} [A + u(\tau_{1}) C, C] d\tau_{1} +$$

$$+ \int_{t_{1}<\tau_{2}<\tau_{1}

$$+ \int_{t_{1}<\tau_{k}<...<\tau_{1}

$$+ O((t - t_{1})^{k+1}) \in L.$$$$$$

If we subtract from the matrix $Y^{-1}(t) CY(t)$ the first k addend, then according to the induction assumption we do not leave L and we get

$$\int_{t_1 < \tau_k < ... < \tau_1 < t} [A + u(\tau_k) C, ..., A + u(\tau_1) C, C] d\tau_k ... d\tau_1 + O((t - t_1)^{k+1}) \in L.$$

Let us multiply the left side by $(t - t_1)^{-k}$ and tend $t \to t_1$, then we get

$$\lceil (A + u(t_1) C)^k C \rceil / k! \in L.$$

We can take another point $t_2 \in \langle 0, T \rangle - G$ for which $u(t_2) \neq u(t_1)$ and by the same procedure we get $[(A + u(t_2) C)^k C] \in L$, k = 0, 1, 2, ... Thus according to lemma 2.1 it holds $\mathfrak{B}(A, B) \subset L$.

2) For t = T we get $X(T) X^{-1}(T) CX(a) X^{-1}(a) = C \in L_1$. Let us, for brevity, denote $X(t_1) X^{-1}(a) = Z$, then we can write $X(T) X^{-1}(t) CX(t) X^{-1}(T) = Z^{-1}Y^{-1}(t) CY(t) Z$, $t \in \{0, T\}$.

Let us choose $t_0 \in \langle 0, T \rangle$, then $[A, Z^{-1}Y^{-1}(t_0) CY(t_0) Z] = [Z^{-1}(ZAZ^{-1}) Z, Z^{-1}Y^{-1}(t_0) CY(t_0) Z] = Z^{-1}[ZAZ^{-1}, Y^{-1}(t_0) CY(t_0)] Z$. According to lemma 1.1 it holds $ZAZ^{-1} \in \mathfrak{A}(A, B)$ and thus in accordance with the first part of this proof we get $[ZAZ^{-1}, Y^{-1}(t_0) CY(t_0)] \in L$, $[A, Z^{-1}Y^{-1}(t_0) CY(t_0) Z] \in L_1$. In the same way we get $[B, Z^{-1}Y^{-1}(t_0) CY(t_0) Z] \in L_1$.

This concludes the proof.

Example. The assumption in lemma 2.8 that the control u is not constant is necessary. Let us put

$$A = \begin{pmatrix} 0 & 0 & 0 \\ -1 & 0 & 0 \\ 0 & -1 & 0 \end{pmatrix}, \quad C = \begin{pmatrix} 2 & 0 & 0 \\ 2 & 0 & 0 \\ 0 & 2 & 0 \end{pmatrix}, \quad u(t) \equiv \frac{1}{2}.$$

We shall show that then (2.8) will not hold for all T > 0.

It holds

$$[A, C] = [A + \frac{1}{2}C, C] = \begin{pmatrix} 0 & 0 & 0 \\ 2 & 0 & 0 \\ 0 & 0 & 0 \end{pmatrix},$$

$$\begin{split} \left[\mathbf{A},\,\mathbf{A},\,\mathbf{C} \right] &= \begin{pmatrix} 0 & 0 & 0 \\ 0 & 0 & 0 \\ 2 & 0 & 0 \end{pmatrix}, \quad \left[\mathbf{A} \,+\, \frac{1}{2}\mathbf{C},\,\mathbf{A} \,+\, \frac{1}{2}\mathbf{C},\,\mathbf{C} \right] = \left[\mathbf{A},\,\mathbf{C} \right], \\ \mathscr{V}(\omega) &= \left\{ \begin{pmatrix} \omega_1 \\ \omega_1 \\ \omega_2 \end{pmatrix}, \, \begin{pmatrix} 0 \\ \omega_1 \\ 0 \end{pmatrix}, \, \begin{pmatrix} 0 \\ 0 \\ \omega_1 \end{pmatrix} \right\}, \\ e^{(\mathbf{A} + \frac{1}{2}\mathbf{C})T} &= \begin{pmatrix} e^T & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1 \end{pmatrix}, \quad \mathscr{V}(e^{(\mathbf{A} + \frac{1}{2}\mathbf{C})T}\omega) &= \left\{ \begin{pmatrix} e^T \omega_1 \\ e^T \omega_1 \\ \omega_2 \end{pmatrix}, \, \begin{pmatrix} 0 \\ e^T \omega_1 \\ 0 \end{pmatrix}, \, \begin{pmatrix} 0 \\ 0 \\ e^T \omega_1 \end{pmatrix} \right\}. \end{split}$$

If $\omega_1 \neq 0$, then for all $T \geq 0$ it holds dim $\mathscr{V}(e^{(A+\frac{1}{2}C)T}\omega) = 3$. For $t \in \langle 0, T \rangle$ we get $X(T)X^{-1}(t)CX(t)X^{-1}(T) = e^{-(A+\frac{1}{2}C)(t-T)}Ce^{(A+\frac{1}{2}C)(t-T)} = C + \sum_{r=1}^{\infty} ((t-T)^r/r!).$ $[(A+\frac{1}{2}C)^rC] = C + (e^{t-T}-1)[A,C].$

Thus it holds

$$\sum_{r=1}^{\infty} r K_{u}(T) = \{C_{\omega}, [A, C] \omega\}, \quad \dim \sum_{r=1}^{\infty} r K_{u}(T) = 2.$$

Lemma 2.9. $T > 0, \ \delta \in (0, \frac{1}{2}), \ u \in M(\delta, 1 - \delta).$ Then

$$(2.9) K_{u}(T) \subset \mathscr{V}(x(T, u)).$$

Proof. According to lemma 2.6 the set $K_u(T)$ is contained in the linear hull L of the vectors (2.7) and from lemma 1.1 it follows $L \subset \mathcal{V}(x(T, u))$.

Lemma 2.10. T > 0, $\delta \in (0, \frac{1}{2})$, $u \in M(\delta, 1 - \delta)$. Let us denote by $\mathcal{T}(x(T, u))$ the set of all possible limits (if they exist),

$$\lim_{k\to\infty} a \frac{x_k - x(T, u)}{\|x_k - x(T, u)\|},$$

where $x_k \in \mathcal{S}_{\omega}(T)$, $x_k \neq x(T, u)$, $k = 1, 2, ..., a \in E_1, x_k \rightarrow x(T, u)$.

Then

$$(2.10) K_{u}(T) \subset \mathscr{T}(x(T, u)).$$

Proof is evident from (2.3).

Lemma 2.11. Let x(t, u) be a solution of (2.1), then dim $\mathscr{V}(x(t, u)) = \dim \mathscr{V}(\omega)$, $t \ge 0$.

Proof. According to theorem 1.1 it holds dim $V(x(t, u)) = \dim V(\omega)$, $t \ge 0$. So it is sufficient to prove the equivalence

$$A\omega \in \mathscr{V}(\omega) \Leftrightarrow Ax(t, u) \in \mathscr{V}(x(t, u))$$

Let us fix $t \ge 0$, then it holds $A\omega \in \mathscr{V}(\omega) \Rightarrow X^{-1}(t) AX(t) \omega \in \mathscr{V}(\omega) \Rightarrow \text{such } A_i \in \mathfrak{B}(A, B)$ and $a_i \in E_1$, i = 1, 2, ..., r, exist that $X^{-1}(t) AX(t) \omega = \sum_{i=1}^r a_i A_i \omega \Rightarrow AX(t, u) = AX(t) \omega = X(t) X^{-1}(t) AX(t) \omega = X(t) \sum_{i=1}^r a_i A_i \omega = (\sum_{i=1}^r a_i X(t) A_i X^{-1}(t)) X(t, u) \Rightarrow AX(t, u) \in \mathscr{V}(X(t, u)).$

The inverse implication we get if we take the matrices -A, -B instead of the matrices A, B.

Theorem 2.1. T > 0, $\omega \in E_u$, dim $\mathscr{V}(\omega) = q$. Then $\mathscr{S}_{\omega}(T)$ is a closed, q-dimensional integral manifold of the distribution \mathscr{V} .

Proof. Let us choose $x \in \mathcal{S}_{\omega}(T)$ and $\varepsilon > 0$. Then there exist $\delta \in (0, \frac{1}{2})$ and a non-constant control $u \in M(\delta, 1 - \delta)$ such that it holds $||x - x(T, u)|| < \varepsilon$.

According to lemma 2.11 dim $\mathcal{V}(x(T, u)) = q$ and from lemma 2.8 it follows that there exist such functions $v_i \in M(-1, 1)$, i = 1, 2, ..., q, that $\mathcal{V}(x(T, u)) = \{x_1(T, v_1), x_1(T, v_2), ..., x_1(T, v_q)\}$, where $x_1(T, v_i)$, i = 1, 2, ..., q, are vectors (2.6).

The function

(2.11)
$$x(T, u + \sum_{i=1}^{q} \vartheta_{i} v_{i}); \quad \vartheta \in G = E\left(\vartheta \in \mathbb{E}_{q}; \|\vartheta\| < \frac{1}{q} \delta\right),$$

represents a mapping of the open set $G \subset E_q$ into $\mathscr{S}_{\omega}(T)$, has continuous partial derivatives of first order with respect to ϑ_i , i = 1, 2, ..., q, which are for $\vartheta = 0$ solutions of the equation

$$\frac{\mathrm{d}}{\mathrm{d}t}\frac{\mathrm{d}x}{\mathrm{d}\theta_i} = \left(\mathbf{A} + u\mathbf{C}\right)\frac{\partial x}{\partial \theta_i} + \mathbf{C}v_i x(t, u); \quad \frac{\partial x}{\partial \theta_i}\Big|_{t=0} = 0; \quad i = 1, 2, ..., q.$$

Hence, the matrix $dx/\partial \vartheta|_{\vartheta=0}$ has as its columns the vectors $x_1(T, \vartheta_i)$, i=1, 2, ..., q, and in an environ of the point $\vartheta=0$ has the rank q.

Thus the set $\mathscr{S}_{\omega}(T)$ is the adherence of a union of a system of q-dimensional integral manifolds of the distribution \mathscr{V} . Let $S_{1,2}$ be two manifolds of this system. Let us choose points $x(T, u_i) \in S_i$, i = 1, 2, where the controls u_i , i = 1, 2 are so chosen that for every $\lambda \in \langle 0, 1 \rangle$ the function $u_1(1 - \lambda) + u_2\lambda$ is not constant. Then the curve

$$(2.12) x(T, u_1(1-\lambda) + u_2\lambda), \quad \lambda \in \langle 0, 1 \rangle,$$

links both points $x(T, u_i)$, i = 1, 2, and each point of (2.12) is contained in a q-dimensional integral manifold of the distribution \mathcal{V} , that is contained in $\mathcal{S}_{\omega}(T)$.

According to theorem 1.1 the theorem is proved.

Theorem 2.2. T > 0, $\omega \in E_n$, dim $V(\omega) = r$. Then $S_{\omega}(T)$ is a closed r-dimensional integral manifold of the distribution V.

Proof. At first let us again show that $S_{\omega}(T)$ is an adherence of a system of r-dimensional integral manifolds of the distribution V. We shall distinguish two cases:

- 1) dim $V(\omega) = \dim \mathscr{V}(\omega)$. Then the statement follows immediately from theorem 2.1.
- 2) dim $V(\omega) > \dim V(\omega)$. Let us choose $x \in S_{\omega}(T)$ and $\varepsilon > 0$, then there exist $\delta \in (0, \frac{1}{2})$, $t \in (0, T)$ and a non-constant continuous on $\langle 0, T \rangle$ control $u \in M(\delta, 1 \delta)$ so that $||x x(t, u)|| < \varepsilon$. If we choose functions $v_i \in M(-1, 1)$, i = 1, 2, ..., r 1, like in theorem 2.1, where we have t instead of T, then the function

(2.13)
$$x(\tau, u + \sum_{i=1}^{r-1} \vartheta_i v_i),$$

$$|\tau - t| < \Delta = \min(t, T - t), \quad \vartheta \in G = E \left[\vartheta \in \mathcal{E}_{r-1}; \|\vartheta\| < \frac{1}{r-1} \delta \right],$$

is a mapping of the open set $(t - \Delta, t + \Delta) \times G \subset E_r$ into $S_{\omega}(T)$. The function (2.13) has all properties as the function (2.11) has and moreover

$$\frac{\mathrm{d}x(\tau, u)}{\mathrm{d}\tau}\bigg|_{\mathfrak{z}=0} = (\mathrm{A} + u(\tau) \mathrm{C}) x(\tau, u).$$

Hence, the functional matrix of the mapping (2.13) has at the point $\tau = t$, $\vartheta = 0$ the rank r.

Now, let us take two integral manifolds $S_{1,2}$ of the distribution V, which are contained in $S_{\omega}(T)$. We choose points $x(t_i, u_i) \in S_i$, i = 1, 2, so that $t_i \in (0, T)$, the control u_i is continuous on $\langle 0, T \rangle$, i = 1, 2, and the function $u_1(1 - \lambda) + u_2\lambda$ is not constant on $\langle 0, \min(t_1, t_2) \rangle$. Let for example be $t_1 < t_2$, then we link both points $x(t_i, u_i)$ with the curve, composed of the following arcs:

$$x\big(t_1,\,u_1\big(1\,-\,\lambda\big)\,+\,u_2\lambda\big)\,,\quad\lambda\in\langle 0,\,1\rangle\,,\quad x\big(t,\,u_2\big)\,,\quad t\in\langle t_1,\,t_2\rangle\,.$$

The theorem is proved.

3. CASE OF 1-DIMENSIONAL MANIFOLD $\mathscr{S}_{\omega}(T)$

In this paragraph we show that every point on a 1-dimensional manifold $\mathscr{S}_{\omega}(T)$ can be reached by a piecewise constant control $u \in M(0, 1)$ and we describe the points at which u is discontinuous. In the whole paragraph we shall have fixed matrices $A, B \in \mathscr{E}_n$.

Lemma 3.1.

(3.1)
$$\frac{\mathrm{d}}{\mathrm{d}u}(A + uC)^{k} = \sum_{i=1}^{k} {k \choose i} (-1)^{i-1} \left[(A + uC)^{i-1} C \right] (A + uC)^{k-i};$$

$$k = 0, 1, 2, \dots$$

Proof by the mathematical induction.

Lemma 3.2.

(3.2)
$$\frac{\mathrm{d}}{\mathrm{d}u} e^{\mathbf{A} + u\mathbf{C}} = \left(\sum_{i \ge 0} \frac{(-1)^i}{(i+1)!} \left[(\mathbf{A} + u\mathbf{C})^i \, \mathbf{C} \right] \right) e^{\mathbf{A} + u\mathbf{C}}.$$

Proof.

$$\frac{d}{du} e^{A+uC} = \sum_{k \ge 0} \frac{1}{k!} \frac{d}{du} (A + uC)^k = \sum_{k \ge 0} \sum_{i=1}^k \frac{1}{k!} {k \choose i} (-1)^{i-1} \left[(A + uC)^{i-1} C \right].$$

$$\cdot (A + uC)^{k-i} = \sum_{i \ge 1} \sum_{s \ge 0} \frac{1}{i! \, s!} (-1)^{i-1} \left[(A + uC)^{i-1} C \right] (A + uC)^s =$$

$$= \sum_{i \ge 0} \frac{(-1)^i}{(i+1)!} \left[(A + uC)^i C \right] e^{A+uC}$$

we use the new index s = k - i.

Lemma 3.3. T > 0, $u \in M(0, 1)$, $\omega \in E_n$, dim $\mathscr{V}(\omega) > 0$. Then the set

$$(3.3) E(t \in \langle 0, T \rangle; Cx(t, u) = 0)$$

is finite.

Proof. Let there exist a sequence $t_1, t_2, t_2, \ldots t_i \in \langle 0, T \rangle$, $t_i \to t_0$, $t_i \neq t_0$, $Cx(t_i, u) = 0$, $i = 1, 2, \ldots$ Let us denote $x(t_0, u) = x_0$. Evidently $Cx_0 = 0$. Let us suppose $CA^rx_0 = 0$, $r = 0, 1, \ldots, (k-1)$. Without loss of generality let us suppose that there exists an infinite number of those terms t_i in the sequence t_1, t_2, \ldots for which $t_i > t_0$. Crossing to a subsequence we can assume that $t_i > t_0$ holds for all $i = 1, 2, \ldots$ Then $t_0 < T$ and we can write for $t \in (t_0, T)$

(3.4)
$$x(t, u) = x_0 + \int_{t_0 < \tau_1 < t} (A + u(\tau_1) C) x_0 d\tau_1 + \dots +$$

$$+ \int_{t_0 < \tau_k < \dots < \tau_1 < t} (A + u(\tau_1) C) \dots (A + u(\tau_k) C) x_0 d\tau_k \dots d\tau_1 +$$

$$+ \int_{t_0 < \tau_{k+1} < \dots < \tau_1 < t} (A + u(\tau_1) C) \dots (A + u(\tau_{k+1}) C) x(\tau_{k+1}, u) d\tau_{k+j} \dots d\tau_1 .$$

If the number of those terms of the sequence $t_1, t_2, ...$, for which $t_i > t_0$ holds, is finite, then $t_0 > 0$ and for $t < t_0$ we would have to change the signs of the inequalities in (3.4) and to alternate the signs of the integrals.

If we put in (3.4) $t = t_i$, i = 1, 2, ..., and multiply it by the matrix

$$\left(\frac{1}{(t_i-t_0)^k}\cdot C\right),$$

we get

$$0 = \frac{1}{(t_i - t_0)^k} Cx(t_i, u) = \frac{1}{(t_i - t_0)^k} \int_{t_0 < \tau_k < \dots < \tau_1 < t_i} CA^k x_0 d\tau_k \dots d\tau_1 +$$

$$+ \frac{1}{(t_i - t_0)^k} \int_{t_0 < \tau_{k+1} < \dots < \tau_1 < t_i} (A + u(\tau_1) C) \dots (A + u(\tau_{k+1}) C) x(\tau_{k+1}, u) d\tau_{k+1} \dots$$

$$\dots d\tau_1 = \frac{1}{k!} CA^k x_0 + O(t_i - t_0).$$

Thus we have found out that it holds $CA^kx_0 = 0$ for all integers k.

Now, let us rewrite the equation (2.1) into the following form:

$$\dot{x} = (A + uC) x = (B + (u - 1) C) x, \quad x(0) = \omega.$$

By the same procedure we would find out that also $CB^kx_0 = 0$, k = 0, 1, 2, ...Hence, according to lemma 2.2 it holds dim $\mathscr{V}(x_0) = 0$ and from lemma 2.11 follows dim $\mathscr{V}(\omega) = 0$. We have got a contradiction.

Lemma 3.4. T > 0, $x \in E_n$, dim $\mathscr{V}(x) = 1$, $u_1 < u_2$. Let us put $g(\theta, t) = e^{(A+\theta C)t}x$, $\theta \in \langle u_1, u_2 \rangle$, $t \in \langle 0, T \rangle$.

Then there exists an analytic function $\psi(\vartheta, t)$, defined on the set

$$G = E(\vartheta \in \langle u_1, u_2 \rangle, t \in \langle 0, T \rangle; \ \mathrm{C}g(\vartheta, t) \neq 0)\,,$$

such that

(3.5)
$$\frac{\partial g(\vartheta, t)}{\partial \vartheta} = \psi(\vartheta, t) \, \mathrm{C}g(\vartheta, t) \,, \quad (\vartheta, t) \in G \,.$$

Moreover, if there exists such $\vartheta_0 \in \langle u_1, u_2 \rangle$ that $Cg(\vartheta_0, t) \neq 0$ for all $t \in (0, T)$, then $\psi(\vartheta_0, t) > 0$ for all $t \in (0, T)$ and there exists $\lim_{t \to 0^+} t^{-1} \cdot \psi(\vartheta_0, t) > 0$.

Proof. In accordance with lemma 3.2 it holds

$$\frac{\partial g(\vartheta, t)}{\partial \vartheta} = \sum_{k \ge 0} (-1)^k \frac{t^{k+1}}{(k+1)!} \left[(A + \vartheta C)^k C \right] g(\vartheta, t).$$

According to lemma 2.11 it holds dim $\mathcal{V}(g(9, t)) = 1$ and thus

$$[(A + \Im C)^k C] g(\Im, t) \in \{Cg(\Im, t)\}, \quad k = 0, 1, 2, ..., (\Im, t) \in G.$$

Let us denote $[(A + \Im C)^k C] = D_k(\Im)$, k = 0, 1, 2, ... As the sequence of norms $||D_k(\Im)||$, k = 0, 1, 2, ... can be majorised by a geometric sequence with the quotient. $2 \max_{\Im \in \langle u_1, u_2 \rangle} ||A + \Im C||$, the first part of the lemma is proved.

It holds:

$$e^{-(A+9C)t} \frac{\partial g}{\partial \vartheta} = e^{-(A+9C)t} \sum_{k \ge 0} (-1)^k \frac{t^{k+1}}{(k+1)!} D_k(\vartheta) e^{(A+9C)t} x =$$

$$= \sum_{k \ge 0} (-1)^k \frac{t^{k+1}}{(k+1)!} e^{-(A+9C)t} D_k(\vartheta) e^{(A+9C)t} x =$$

$$= \sum_{k \ge 0} (-1)^k \frac{t^{k+1}}{(k+1)!} \left(\sum_{e \ge 0} \frac{t^e}{e!} D_{k+e}(\vartheta) \right) x = \sum_{r \ge 0} \frac{t^{r+1}}{(r+1)!} D_r(\vartheta) x ,$$

$$e^{-(A+9C)t} \frac{\partial g}{\partial \vartheta} = e^{-(A+9C)t} \psi(\vartheta, t) Cg =$$

$$= \psi(\vartheta, t) e^{-(A+9C)t} Ce^{(A+9C)t} x = \psi(\vartheta, t) \sum_{s \ge 0} \frac{t^r}{r!} D_r(\vartheta) x .$$

For $\vartheta = \vartheta_0$ and for sufficiently small t > 0 both sides of (3.5) are not zero. Let us denote by $\beta(t)$ one non-zero coordinate of the vector $e^{-(A+\vartheta_0C)t}$. $(\partial g/\partial \vartheta)$, then we get the equation

$$\beta(t) = \psi(\vartheta_0, t) \cdot \beta'(t), \quad t \in (0, t_0), \quad t_0 = \min \left(T, \inf_{\substack{\beta(\tau) = 0 \\ \tau > 0}} \tau\right).$$

Let be $D_i(\theta_0) x = 0$; i = 0, 1, ..., r - 1; $D_r(\theta_0) x \neq 0$, then evidently $\lim_{t \to 0+} t^{-1}$. $\psi(\theta_0, t) = 1/(r + 1) > 0$.

Let us take $t_1 \in (0, t_0)$ so small that $\psi(\vartheta_0, t_1) > 0$, then

$$\beta(t) = \beta(t_1) \exp \left\{ \int_{t_1}^t (1/\psi(\vartheta_0, t)) dt \right\}, \quad t \in \langle t_1, t_2 \rangle, \quad t_2 = \min(T, \inf_{\beta'(\tau) = 0} \tau \rangle_{t_1}$$

It is not possible to be $t_0 < t_2$ as for $t \in \langle t_1, t_2 \rangle$ it holds $|\beta(t)| \ge |\beta(t_1)| > 0$.

If it holds $t_2 < t_0$, then there is $\lim_{t \to t_2 -} \psi(\theta_0, t) = +\infty$, what is again impossible, as

 $\psi(\vartheta_0, t)$ is bounded on $\langle t_1, T \rangle$. Thus it holds $t_0 = t_2$.

If it holds $t_2 < T$, then another coordinate $\tilde{\beta}$ of the vector $\exp \{-(A + \vartheta_0 C) t\}$. $(\partial g/\partial \vartheta)$ has its derivative at the point t_2 different from zero and again it holds

$$\tilde{\beta}(t) = \psi(\vartheta_0, t) \cdot \tilde{\beta}'(t), \quad t \in \langle t_2 - \Delta, t_3 \rangle, \quad t_3 = \min \left(T, \inf_{\substack{\tilde{\beta}'(\tau) = 0 \\ \tau > t_2}} \tau \right),$$

where $\Delta > 0$ is so small that $\tilde{\beta}'(t) \neq 0$ on the interval $\langle t_2 - \Delta, t_2 \rangle$.

This completes the proof.

Lemma 3.5.
$$u_{1,2} \in \langle 0, 1 \rangle$$
, $t_{1,2} > 0$, $\omega \in E_n$, dim $\mathscr{V}(\omega) = 1$. Let us put
$$x = e^{(A + u_2 C)t_2} e^{(A + u_1 C)t_1} \omega$$
.

Then it exists a piecewise constant control

$$u \in M(\min(u_1, u_2), \max(u_1, u_2))$$

such that:

- 1) $x = x(t_1 + t_2, u),$
- 2) if t is a point of discontinuity of the control u, then Cx(t, u) = 0.

Proof. The case $u_1 = u_2$ is trivial. Let be $u_1 < u_2$. Let us denote by u_0 the control given by the prescription:

$$u_0(t) = u_1$$

$$u_0(t) = u_2$$
for
$$\begin{cases} t \in \langle 0, t_1 \rangle \\ t \in \langle t_1, t_1 + t_2 \rangle \end{cases}$$

If it is $t_1 \in F = E(t \in (0, t_1 + t_2); Cx(t, u_0) = 0)$, then there is nothing to be proved. So let be $t_1 \notin F$. Without loss of generality we can assume that $F = \emptyset$. Then on some open neighbourhood G of the set of all points $x(t, u_0)$, where $t \in (0, t_1 + t_2)$, it holds $x \in G \Rightarrow Cx \neq 0$.

Let us choose $\Delta \in (0, t_2)$ so small that for all $\Delta_{1,2} \ge 0$, $\Delta_1 + \Delta_2 \le \Delta$, it holds

$$e^{(\mathbf{A}+u_2\mathbf{C})\Delta_2} e^{(\mathbf{A}+u_1\mathbf{C})(t_1+\Delta_1)} \omega \in G.$$

Let us put

$$f(\tau,\,\zeta) = e^{(\mathbf{A} + u_2 C)\tau} \, e^{(\mathbf{A} + u_1 C)(t_1 + \zeta - \tau)} \, \omega \; ; \quad \tau,\,\zeta \in \langle 0,\,\Delta \rangle \; ,$$

Then it holds:

(3.6)
$$\frac{\partial f(\tau,\zeta)}{\partial \tau} = (u_2 - u_1) e^{(A + u_2 C)\tau} C e^{(A + u_1 C)(\tau_1 + \zeta - \tau)} \omega =$$
$$= (u_2 - u_1) e^{(A + u_2 C)\tau} C e^{-(A + u_2 C)\tau} f(\tau,\zeta).$$

The right side of (3.6) is for $\tau = 0$ different from zero. We can choose Δ so small that $\partial f(\tau, \zeta)/\partial \tau = 0$ for $\tau, \zeta \in \langle 0, \Delta \rangle$.

Let us further put

$$g\big(\vartheta,\,t\big)=\,e^{(\mathsf{A}\,+\,\vartheta\mathsf{C})t}\omega\;,\quad\vartheta\in\left\langle u_1,\,u_2\right\rangle\;,\quad t\in\left\langle 0,\,t_1\,+\,t_2\right\rangle\;.$$

According to lemma 3.4 we can define functions φ , ψ as follows:

$$\frac{\partial f(\tau,\zeta)}{\partial \tau} = \varphi(\tau,\zeta) \, \mathrm{C} f(\tau,\zeta) \,, \quad \frac{\partial g(\vartheta,t)}{\partial \vartheta} = \psi(\vartheta,t) \, \mathrm{C} g(\vartheta,t) \,.$$

Then φ is analytic and different from zero on the set τ , $\zeta \in \langle 0, \Delta \rangle$. As $\varphi(0, \zeta) = u_2 - u_1 > 0$, $\zeta \in \langle 0, \Delta \rangle$, the function φ is positive. The function ψ is according to lemma 3.4 analytic on the set of all (9, t) for which $Cg(9, t) \neq 0$ and according to the supplement to lemma 3.4 it is $\psi(u_1, t) > 0$ for $t \in (0, t_1 + \Delta)$.

Let us put $K_1 = \max \varphi(\tau, \zeta)$, where $\tau, \zeta \in \langle 0, \Delta \rangle$ and choose $\Delta_1 > 0$ so small that $\psi(\vartheta, t) > 0$ for $\vartheta \in \langle u_1, u_1 + \Delta_1 \rangle$, $t \in \langle t_1, t_1 + \Delta \rangle$. Let us put $K_2 = \min \psi(\vartheta, t)$, where $\vartheta \in \langle u_1, u_1 + \Delta_1 \rangle$, $t \in \langle t_1, t_1 + \Delta \rangle$, $\tau_0 = \min \left(\Delta, \left(K_2/K_1\right)\Delta_1, \left(K_2/K_1\right)\left(u_2 - u_1\right)\right)$ and take the equation

$$\frac{\mathrm{d}\vartheta}{\mathrm{d}\tau} = \frac{\varphi(\tau,\zeta)}{\psi(\vartheta,t_1+\zeta)}, \quad \vartheta(0,\zeta) = u_1, \quad \tau \in \langle 0,\tau_0 \rangle, \quad \zeta \in \langle 0,\Delta \rangle.$$

The function φ is defined on $\langle 0, \tau_0 \rangle$ and it holds

$$u_1 \leq \vartheta(\tau,\zeta) \leq u_1 + \int_0^{\tau} \frac{\varphi(\tau,\zeta)}{\psi(\vartheta,t_1+\zeta)} d\tau \leq u_1 + \frac{K_1}{K_2} \tau \leq \min(u_1+\Delta_1,u_2).$$

Hence, the solution ϑ exists on the interval $\langle 0, \tau_0 \rangle$ for $\zeta \in \langle 0, \Delta \rangle$.

Let us define the function $h(\tau) = g(\vartheta(\tau, \tau_0), t_1 + \tau_0), \tau \in \langle 0, \tau_0 \rangle$. Then it holds

$$\frac{\mathrm{d}h(\tau)}{\mathrm{d}\tau} = \frac{\partial g(\vartheta(\tau, \tau_0), t_1 + \tau_0)}{\partial \vartheta} \cdot \frac{\partial \vartheta(\tau, \tau_0)}{\partial \tau} = \varphi(\tau, \tau_0) \, \mathrm{C}g(\vartheta(\tau, \tau_0), t_1 + \tau_0) =$$

$$= \varphi(\tau, \tau_0) \, \mathrm{C}h(\tau) \quad \text{for} \quad \tau \in \langle 0, \tau_0 \rangle \, .$$

$$h(0) = g(\Im(0, \tau_0), t_1 + \tau_0) = g(u_1, t_1 + \tau_0) = e^{(\mathbf{A} + u_1 C)(t_1 + \tau_0)} \omega = f(0, \tau_0).$$

Thus we have got $h(\tau) = f(\tau, \tau_0)$ for $\tau \in \langle 0, \tau_0 \rangle$. If we put $\tau = \tau_0$, we get $f(\tau_0, \tau_0) = g(\vartheta(\tau_0, \tau_0), t_1 + \tau_0)$ i.e.

$$e^{(A + u_2 C) \tau_0} \; e^{(A + u_1 C) t_1} \; \omega \; = \; e^{(A + \vartheta(\tau_0, \tau_0) C) (t_1 + \tau_0)} \; \omega \; .$$

If it is $\vartheta(\tau_0, \tau_0) = u_2$, then the proof is finished. If it is $\vartheta(\tau_0, \tau_0) < u_2$ and if the set

$$E(t \in (0, t_1 + \tau_0); Ce^{(A+\vartheta(\tau_0,\tau_0)C)t} \omega = 0)$$

is empty, we get the original problem and we can repeat the whole procedure. So we get (finite or infinite) sequences (let us write only the case of infinite sequences):

$$u_1 < \vartheta_0 < \vartheta_1 < \vartheta_2 \dots < u_2, \quad \tau_0, \tau_1, \tau_2, \dots, ; \quad \tau_i > 0, \quad i = 0, 1, 2, \dots$$

such that either $\sum_{i\geq 0} \tau_i = t_2$, then $x = e^{(A+\tilde{\mathfrak{g}}C)(t_1+t_2)}\omega$, where $\tilde{\mathfrak{g}} = \lim_{i\to\infty} \vartheta_i$ or $\sum_{i\geq 0} \tau_i = \tilde{\tau} < t_2$. If in the second case the set

$$F_1 = E(t \in (0, t_1 + \tilde{\tau}); Ce^{(A + \tilde{s}C)t}\omega = 0)$$

is empty, we could use our procedure in the case

$$x = e^{(A + u_2C)(t_2 - \tilde{t})} e^{(A + \tilde{s}C)(t_1 + \tilde{t})} \omega.$$

If $F_1 \neq \emptyset$, then there exists $\tilde{t} = \max_{t \in F_1} t$ (F_1 is according to lemma 3.3 finite). Then we define $u(t) = \tilde{\mathfrak{I}}$ for $t \in \langle 0, \tilde{t} \rangle$ and we get the original problem:

$$x = e^{(\mathbf{A} + u_2 \mathbf{C})(t_2 - \tilde{\tau})} e^{(\mathbf{A} + \tilde{\mathbf{g}} \mathbf{C})(t_1 + \tilde{\tau} - t)} x(\tilde{t}, u).$$

According to lemma 3.3 we must reach the point x after finite number of such steps and the proof is finished.

Theorem 3.1. $\omega \in E_n$, dim $\mathscr{V}(\omega) = 1$, T > 0, $x \in \mathscr{S}_{\omega}(T)$. Then such piecewise constant control $u \in M(0, 1)$ exists that it holds:

- 1) x = x(T, u),
- 2) u has a finite number of discontinuities. Moreover, if $t \in \langle 0, T \rangle$ is a point of discontinuity of u, then Cx(t, u) = 0.

Proof. It exists such $v \in M(0, 1)$ that x = x(T, v). Let $v_k \in M(0, 1)$, k = 1, 2, ... be a sequence of piecewise constant on $\langle 0, T \rangle$ controls such that $v_k \to v$ asymptotically on $\langle 0, T \rangle$. According to lemma 3.5 for each integer k it exists a piecewise constant on $\langle 0, T \rangle$ control w_k that has only such discontinuity-points t at which $Cx(t, \omega_k) = 0$ and it holds $x(T, w_k) = x(T, v_k)$.

We can choose from the sequence w_k , k = 1, 2, ..., such subsequence (let it be the original sequence) that converges asymptotically to a control w. Then according to lemma 2.7 it holds

(3.7)
$$x(t, w_k) \to x(t, w)$$
 uniformly on $\langle 0, T \rangle$.

According to lemma 3.3 the set $E(t \in \langle 0, T \rangle; Cx(t, w) = 0)$ is finite. Let us denote its elements by t_i , i = 1, 2, ..., r, $t_1 < t_2 < t_3 ... < t_r$.

Let us choose $\varepsilon \in (0, \frac{1}{2}(t_2 - t_1))$ and put $\varrho = \min ||y - x(t, w)||$, where $y \in \varepsilon E(x \in E_n; Cx = 0)$, $t \in \langle t_1 + \varepsilon, t_2 - \varepsilon \rangle$.

Evidently $\varrho > 0$. According to (3.7) it exists such integer k_0 that for all $k > k_0$ it holds $||x(t, w_k) - x(t, w)|| < \varrho$, where $t \in \langle t_1 + \varepsilon, t_2 - \varepsilon \rangle$. Thus the control w_k , $k > k_0$, is on $\langle t_1 + \varepsilon, t_2 - \varepsilon \rangle$ constant. The limit w must be also constant on

 $\langle t_1 + \varepsilon, t_2 - \varepsilon \rangle$ for all $\varepsilon \in (0, \frac{1}{2}(t_2 - t_1))$. Hence, w is constant on (t_1, t_2) , what was to be proved.

Theorem 3.2. $\omega \in E_n$, dim $\mathscr{V}(\omega) = 1$, T > 0, $x \in \mathscr{S}_{\omega}(T)$. Let the matrix C be regular. Then there exists such constant control $u \in M(0, 1)$ that x = x(T, u).

The proof follows immediately from theorem 3.1, as the solution of the equation Cx = 0 is just only x = 0.

Example. Let us take

$$A = \begin{pmatrix} 1 & \frac{1}{2} & 1 \\ 0 & 1 & 0 \\ 0 & 0 & 1 \end{pmatrix}, \quad C = \begin{pmatrix} 2 & \gamma_2 & \gamma_3 \\ 0 & 0 & 0 \\ 0 & 0 & 1 \end{pmatrix}, \quad \omega \in E_3, \ \omega_3 = 0$$

and for $t \in \langle 0, T \rangle$ put:

$$f(t) = e^{\mathbf{A}(T-t)} e^{(\mathbf{A}+\mathbf{C})t} \omega = e^{T} \begin{pmatrix} e^{2t} \left(\omega_{1} + \frac{2\gamma_{2}+1}{4} \omega_{2} \right) + \frac{(T-t)2 - (2\gamma_{2}+1)}{4} \omega_{2} \\ \omega_{2} \\ 0 \end{pmatrix}$$

$$g(t) = e^{(\mathbf{A} + (t/T)\mathbf{C})T} \omega = e^{T} \begin{pmatrix} e^{2t}\omega_1 + (e^{2t} - 1)\frac{2t\gamma_2 + T}{4t}\omega_2 \\ \omega_2 \\ 0 \end{pmatrix}.$$

Evidently f(0) = g(0), f(T) = g(T).

If we take such ω that satisfies the inequalities:

$$2\omega_1 + \gamma_2\omega_2 < 0$$
, $4(2\omega_1 + \gamma_2\omega_2) + \omega_2 > 0$,

then it holds dim $\mathscr{V}(\omega) = 1$ and for T > 1 the first coordinate of the vector f, resp. g, at first decreases and then increases, resp. still increases, on $\langle 0, T \rangle$.

Hence, we cannot reach every point of $\mathcal{S}_{\omega}(T)$, where T > 1, by a constant control at the time T.

References

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Author's address: Praha 1, Žitná 25, ČSSR (Matematický ústav ČSAV).

Резюме

РЕШЕНИЕ В ЦЕЛОМ УРАВНЕНИЯ УПРАВЛЕНИЯ

 $\dot{x} = (A(1-u) + Bu) x$

ЯН КУЧЕРА (Jan Kučera), Прага

В данной работе исследуется множество $\mathcal{S}_{\omega}(T)$ или $S_{\omega}(T)$ всех точек, в которые возможно попасть из данной начальной точки ω по решению уравнения (2.1) за данное время T или во время менше или равно T. Доказано в теоремах 2.1 и 2.2 что эти множества являются замыканием некоторых многообразий, которые локально заданы отображениями (2.11) и (2.13) или отображением (1.2). Размерность этих многообразий равняется размерности некоторого распределения, которое введено в [2], в начальной точке ω .

В третьем параграфе изучается более подробно случай когда $\mathcal{S}_{\omega}(T)$ кривая. Потом в каждую из точек $\mathcal{S}_{\omega}(T)$ можно попасть при помощи по частьях постоянного управления u, точки перерыва которого соответствуют пересечениям x(t,u) с пространством решений уравнения (A-B)x=0.

an Makasa Maraji kangkan menjilikan alabagan Kandiga Senarakan an Bagidiga salah digi k**abigiga** kancalah ali

en et l'élie : l'oright tripare : manière par l'illing : le pas eau conserve : le le pareir : le le le light t The tripare par light d'unit part conserve : l'origan propriét passeur tre en despassion à l'origan de la care Tarrès light i de l'entre de l'illinguation de la propriét de l'interes passeur la sevant au confirme de l'enc