# ON THE UNIQUENESS OF SOLUTION AND CERTAIN APPROXIMATE METHODS 

# OF SOLVING LINEAR VISCOELASTICITY PROBLEMS 

PMM Vol. 37, N83, 1973, pp. 505-514<br>L. P. LEBEDEV<br>(Rostov-on-Don)<br>(Received June 16, 1972)

The uniqueness of solution of mixed initial-boundary value problems for a version of linear viscoelasticity is proved, as is also the convergence of some approximate methods of solving these problems. Upon compliance with the conditions $\mu>0, m_{A} \leqslant m_{B}$ and if $m_{A}=m_{B}$, then $\lambda+2 \mu>0$ the following cases of the relationship between powers of the polynomials $B$ and $C$ are considered: $m_{B} \geqslant m_{C}+2$ (Sect. 3), $m_{B}=m_{C}$ and $m_{B}=m_{C}+1$ (Sect. 4). These cases include all the types of relations between the strains and stresses of the type(1.1) used in the theory of viscoelasticity. As a mathematical foundation for such constraints, it can be noted that for $m_{C}>m_{B}$ the Cauchy problem for a body occupying all of space is always incorrect according to G. E. Shilov [1].

1. A version of linear viscoelasticity is considered with the following governing relationships between the stress $\sigma_{i j}$ and strain $\varepsilon_{i j}$ tensors:

$$
\begin{gather*}
C\left(\partial_{l}\right) \sigma_{i j}=\lambda A\left(\partial_{t}\right) \theta \delta_{i j}+2 \mu B\left(\partial_{t}\right) \varepsilon_{i j}  \tag{1.1}\\
\varepsilon_{i j}=1 / 2\left(u_{i, j}+u_{j, i}\right)
\end{gather*}
$$

Here $A(p), B(p)$ and $C(p)$ are some polynomials of the variable $p$ with the highest coefficient equal to unity and the powers $m_{A}, m_{B}=n, m_{C}=m-2$, respectively, $\lambda$ and $\mu$ are the instantaneous elastic moduli, $\theta=\varepsilon_{i t}$ is the volume strain, $u_{i, j}$ is the partial derivative of the $i$ th component of the displacement vector $\mathbf{u}$ of points of the body with respect to the variable $x_{j}, \partial_{t}$ is the partial derivative with respect to time $t$, and $\delta_{i j}$ is the Kronecker symbol; summation is performed over the repeated subscripts.

It is natural to use the principle of possible displacements in the Lagrange form

$$
\begin{equation*}
\int_{\Omega} \sigma_{i j} \delta \varepsilon_{i j} d \Omega=\int_{\Omega}\left(F_{i}-\rho \partial_{t}^{2} u_{i}\right) \delta u_{i} d \Omega+\int_{S} P_{i} \delta u_{i} d S \tag{1.2}
\end{equation*}
$$

for a generalized formulation of viscoelasticitv theorv problems. where $\delta u_{i}$ is the variation of the displacement vector components, $\mathbf{F}=\left(F_{1}, F_{2}, F_{3}\right)$ are the volume forces, $\mathbf{P}=\left(P_{1}, P_{2}, P_{3}\right)$ are forces acting on the boundary $S$ of the volume $\Omega$ occupied by the body, and $\rho$ is the body density. Considering the variations $\delta u_{i}$ independent of time $t$, let us eliminate $\sigma_{i j}$ from (1.2) by applying the operator $C\left(\partial_{i}\right)$ to (1.2) and taking account of (1.1). If the vector function $u$ in the equality thus obtained is a vector of the actual displacements of the body points, then this equality must be satisfied at any time $t$ and for all arbitrary possible displacements $\delta u_{i}$, and it hence continues to be satisfied if $\delta u_{i}$ depend on time $t$. Let us present this equality which we have additionally integrated with respect to time

$$
\begin{gather*}
\int_{0}^{T} \int_{\Omega}\left\{\lambda A\left(\partial_{t}\right) \theta \delta_{i j}+2 \mu B\left(\partial_{t}\right) \varepsilon_{i j}\right\} \delta \varepsilon_{i j} d \Omega d t= \\
\int_{0}^{T} \int_{\Omega} \delta u_{i} C\left(\partial_{i}\right)\left(F_{i}-\rho \partial_{i}^{2} u_{i}\right) d \Omega d t+\int_{0}^{T} \int_{S} \delta u_{i} C\left(\partial_{t}\right) P_{i} d S d t \tag{1.3}
\end{gather*}
$$

The equality (1.3) is taken as the basis in the generalized formulation of linear viscoelasticity problems below.

For a complete formulation of the problem it is necessary to give the following initial and boundary conditions:

$$
\begin{gather*}
\left.\partial_{t}{ }^{k} \mathbf{u}\right|_{t=0}=\mathbf{a}_{n}, \quad k=0, \ldots, \max (n-1, m-1)  \tag{1.4}\\
\left.N_{1} \mathbf{u}\right|_{S}=N_{\mathbf{1}} \mathbf{\psi}  \tag{1.5}\\
\left.N_{2} \sigma_{k l} v_{l} \mathbf{i}_{k}\right|_{S}=N_{2} \mathbf{P}=\mathbf{P} \tag{1.6}
\end{gather*}
$$

where $v_{l}$ are direction cosines of the exterior normal to the surface $S$, and $N_{1}, N_{2}$ are mutually orthogonal projection operators in the space $E^{3}$ which depend piecewise-continuously on the coordinates of the surface $S$. The assignment of displacements or stresses on the body boundary, respectively, denotes the case of identical operators $N_{1}$ or $N_{2}$

It is furthermore considered that there is a degenerate part $S_{1}$ of the boundary $S$ where the operator $N_{1}$ is identical, i. e. the possibility of body displacement as a rigid whole is eliminated, and the operator $N_{1}+N_{2}$ is identical for all points of $S$.

The following condition must be satisfied.
Consistency condition. There exists a vector function $\boldsymbol{\Phi}$ satisfying the initial and boundary conditions (1.4), (1.5).
2. Let us introduce certain spaces and let us examine their properties. The spaces $C^{k}[0, T ; H], L_{p}[0, T ; H], W_{p}{ }^{k}(\Omega)$ are considered known [2]. That the vector function belongs to any of them means that each of its components belongs to this space.

The boundary $S$ of the bounded volume $\Omega$ possesses a normal varying piecewisecontinuously on $S$ and such that some cone of positive aperture can be adjoined from within at each point $S$.

Definition 2.1. The space $B_{1}^{n}, m(T)$ is the closure of the set of vector-functions $\mathbf{a}=\left(a_{1}, a_{2}, a_{3}\right), \mathbf{a} \in C^{\infty}(\Omega \times[0, T])$ in the norm

$$
\begin{equation*}
\|\mathbf{a}\|_{B_{1}^{n, m}(T)}^{2}=\int_{0}^{T}\left\{\sum_{k=0}^{n}\left\|\partial_{l}^{k} \mathbf{a}\right\|_{W_{2^{2}}(\Omega)}^{2}+\sum_{k=0}^{m}\left\|\partial_{l}^{k} \mathbf{a}\right\|_{L_{2}(\Omega)}^{2}\right\} d t \tag{2.1}
\end{equation*}
$$

Let us introduce the following notation:

$$
\begin{equation*}
(\mathbf{a} \cdot \mathbf{b})_{H_{\mathbf{1}}}=\int_{\Omega}\left\{\lambda \theta(\mathbf{a}) \delta_{i j}+2 \mu \varepsilon_{i j}(\mathbf{a})\right\} \varepsilon_{i j}(\mathbf{b}) d \Omega \tag{2.2}
\end{equation*}
$$

Definition 2.2.The closure of a subset of vector functions $a \in C^{1}(\Omega)$ satisfying the homogeneous boundary conditions (1.5) is the space $H_{1}$ in the norm (2.2).
Definition 2.3. The closure of the subset of vector functions $\mathbf{a} \equiv C^{\infty}(\Omega \times$ $[0, T])$ satisfying the homogeneous boundary and initial conditions (1.4),(1.5) in the norm

$$
\begin{equation*}
\|\mathbf{a}\|_{E_{2}^{n}(T)}^{2}=\int_{0}^{T}\left(\partial_{t}{ }^{n} \mathbf{a} \cdot \partial_{t}{ }^{n} \mathbf{a}\right)_{H_{1}} d t \tag{2.3}
\end{equation*}
$$

is called the space $B_{2}{ }^{n}(T)$ if $\mu>0, \lambda+2 \mu>0$. The space $B_{2}{ }^{\circ}(T)$ has the spectal notation $H_{2}(T)$.

Note. When $S_{1}=S$, the condition $\lambda+2 \mu>0$ can be replaced by the weaker condition $3 \lambda+2 \mu>0$.

Definition 2.4. The closure of the subset of vector-functions $\mathbf{a} \in C^{\infty}(\Omega \times$ $[0, T]$ ) satisfying the homogeneous boundary conditions (1.5) in the norm

$$
\begin{equation*}
\|\mathbf{a}\|_{H_{3}(T)}^{2}=\int_{0}^{T}\left\{(\mathbf{a} \cdot \mathbf{a})_{H_{1}}+\left\|\partial_{l} \mathbf{a}\right\|_{L_{2}(\Omega)}^{2}\right\} d t \tag{2.4}
\end{equation*}
$$

is the space $H_{3}(T)$. The subset $H_{3}(T)$ obtained by the closure of vector functions which equal zero in some neighborhood of $t=T$ for each function is denoted by $H_{3}{ }^{\circ}(T)$.

Definition 2.5. $\quad B_{3}{ }^{n}(T)$ is the subset of vector functions $a \in B_{1}^{n, n}(T)$ such that $\left\|\partial_{t} a\right\|_{L_{z}(\Omega)}$ is a function bounded in $[0, T]$.

Definition 2.6. The space of elements $\mathbf{b}$ dual to the space $H$ and with the norm

$$
\|\mathbf{b}\|_{H^{-}}=\sup \left|(\mathbf{b} \cdot \mathbf{a})_{L_{\mathbf{s}}(\Omega)}\right|\|\mathbf{a}\|_{H}^{-1}, \quad \mathbf{a} \in H
$$

is called the negative space $H^{-}$.
Lemma 2.1. The following imbeddings of spaces with continuous imbedding operator hold:

$$
\begin{gathered}
H_{1} \subset W_{2}^{1}(\Omega), \quad L_{8 / \%}(\Omega) \subset H_{1}{ }^{-} \\
B_{2}^{n}(T) \subset B_{1}^{n, n}(T), \quad B_{1}^{n, m}(T) \subset C^{n-1}\left[0, T ; W_{2}{ }^{1}(\Omega)\right] \\
B_{1}^{n, m}(T) \subset C^{k}\left[0, T ; W_{2}^{1-(k+0.5-n)(m-n)^{-1}}(\Omega)\right] \quad \text { for } n \leqslant k<m \\
B_{1}^{n, m}(T) \subset W_{2}^{n}\left[0, T ; W_{2}^{1 / 2}(\Gamma)\right], \quad H_{3}(T) \subset W_{2}^{1}(\Omega \times[0, T])
\end{gathered}
$$

where the norm (2.2) and the norm $W_{2}{ }^{1}(\Omega)$ are equivalent in the space $H_{1}$, and the norms (2.3) and $B_{1}^{n, n}(T)$ in $B_{2}{ }^{n}(T) ; \quad \Gamma \in \Lambda_{1}$. Here $\Lambda_{1}$ is the class of Liapunov surfaces $\Gamma$, i. e. a finite number of domains $\Gamma_{k}$ each of which is representable parametrically as $z_{k}=f_{k}\left(x_{k}, y_{k}\right)$, where $f_{k}$ are continuously differentiable functions can be isolated on the surface $\Gamma$, where any interior part of the surface $\Gamma$ belongs to the union of some strictly interior closed subdomains $\Gamma_{k}$ for each such part. The proof results from Theorem 3.2 (see [2], Chapt, 1) and the Korn inequality.

Lemma 2.2. Let there be a set of $m$ vector functions. $a_{0}, \ldots, a_{m}$ with the following properties: $\quad \mathbf{a}_{k} \in W_{2}{ }^{1}(\Omega), \quad k=0, \ldots, n-1$

$$
\mathbf{a}_{k} \in W_{2}^{1-(k+0.5-n)(m-n)^{-1}}(\Omega), \quad k=n, \ldots, m-1, \quad \text { if } \quad m>n
$$

In this case there exists a vector function $\mathbf{a} \in B_{1}^{n, m}(T)$ taking this set as the initial values (1.4). If the functions $a_{k}$ satisfy the homogeneous boundary conditions (1.5) in the sense of the spaces to which they belong, then the vector function a also satisfies the homogeneous boundary conditions (1.5). The proof results from Theorem 3.2 (see [2], Chapt. 1).

Lemma 2.3. For any vector function $\omega$ given on some surface $\Gamma \subset \Omega, \Gamma \in \Lambda_{1}$, such that $\partial_{t}{ }^{k} \omega \in L_{2}\left[0, T ; W_{2}^{t_{i}}(\Gamma)\right] k=0, \ldots, n-1$, there exists a vector function $\mathbf{a} \in B_{1}^{n, n}(T)$ taking the value $\omega$ on $\Gamma$, where if $\partial_{t}{ }^{k} \omega=0, k=0, \ldots$, $n-1$ for $t=0$, then $\partial_{t}{ }^{k} \mathbf{a}=0$ also for $t=0$ and the reconstruction operator is
continuous. The proof is carried out analogously to [3].
Lemmas 2.2 and 2.3 indicate important particular cases of the existence of the vector function $\Phi$ from the consistency condition for a definite class. It must be noted that the representation

$$
\begin{equation*}
\mathbf{a}\left(x_{i}, t\right)=\int_{0}^{t} \frac{(t-\tau)^{n-1}}{(n-1)!} \partial_{\tau}^{n} \mathbf{a}\left(x_{i}, \tau\right) d \tau \tag{2.5}
\end{equation*}
$$

is valid for a vector function $\mathbf{a} \in B_{1}^{n, m}(T)$ with initial values $\partial_{t}{ }^{k} \mathbf{a}=0, k=0, \ldots$, $n-1$ for $t=0$.
8. Each of the three cases of the relationship between the powers of the polynomials $B$ and $C$ mentioned in Sect. 1 in connection with the singularities in the formulation of the problem, should be examined separately.
Definition 3.1. The vector function $\mathbf{u} \in B_{1}^{n, n}(T)$ satisfying the initial and boundary conditions (1.4), (1.5) such that the equality (1.3) is valid for any vector function $\delta \mathbf{u} \in H_{2}(T)$ is called the generalized solution of the linear viscoelasticity problem in the case $n=m_{B} \geqslant m_{\mathrm{C}}+2$.

The case $m_{A}=m_{B}$ is examined for definiteness, Let the vector function $\boldsymbol{\Phi}$ from the consistency condition be such that $\boldsymbol{\Phi} \in B_{1}^{n, n}(T)$. We seek the generalized solution of the linear viscoelasticity problem in the form

$$
\begin{equation*}
\mathbf{u}=\mathbf{u}_{0}+\boldsymbol{\Phi} \tag{3.1}
\end{equation*}
$$

where the vector function $\mathbf{u}_{0}$ evidently belongs to the space $B_{2}^{n}(T)$. By introducing the notation $\mathbf{v}=\partial_{t}{ }^{n} \mathbf{u}_{0}$ and taking account of (2.5), (3.1), Eq. (1.3) can be reduced to the following form:

$$
\begin{equation*}
(\mathbf{v} \cdot \delta \mathbf{u})_{H_{2}(T)}=\int_{0}^{T}\left\{R_{1}(\mathbf{v}, \delta \mathbf{u})+R_{2}(\Phi, \mathbf{F}, \mathbf{P}, \delta \mathbf{u})\right\} d t \tag{3.2}
\end{equation*}
$$

where $R_{k}$ is some integro-differential form which is linear in each of its variables. It must be noted that the form $R_{1}(\mathbf{v}, \delta \mathbf{u})$ contains the variable $\mathbf{v}$ only under the integral with respect to time $t$, with the exception of the term $\rho(\mathrm{v} \cdot \delta \mathrm{u})_{L_{z}(\Omega)}$ for $m_{C}+2=$ $m_{B}$, which must be added to the expression for the norm $H_{1}$ in this case. For $\rho>0$ such a new norm $H_{1}$ is equivalent to the old one.

It can be shown that the uniqueness of solution of (3.2) in the space $H_{2}(T)$ is equivalent to the existence of a single generalized solution of the linear viscoelasticity problem. All the properties and methods of the solution of (3.2) are easily carried over directly to (1.3). Therefore, (3.2) will be investigated below, and the results again formulated for (1.3). The investigation is carried out analogously in Sect. 4.

Theorem 3.1. Let the following conditions be satisfied:

1) $\left.\left.m_{C}+2 \leqslant m_{B}=n, \quad m_{A} \leqslant m_{B}, 2\right) \mu>0,3\right)$ if $m_{A}=m_{B}$,
then $\lambda+2 \mu>0, ~ 4)$ if $m_{B}=m_{C}+2$, then $\rho>0$, 5) $\Phi \rightleftharpoons B_{1}^{n, n}(T)$, 6) $\left.C\left(\partial_{t}\right) \mathbf{F} \in L_{2}\left[0, T ; H_{1}^{-}\right], 7\right) C\left(\partial_{t}\right) P_{i} \in L_{2}\left[0, T ; W_{2}^{-1 / 2}(S)\right]$

In this case: 1) there exists a single generalized solution of the linear viscoelasticity problem in the sense that has been introduced by Definition 3.1 ; 2) the solution can be sought approximately by the Bubnov-Galerkin method in the form $\mathbf{u}_{n}{ }^{*}=c_{1}(t) \psi_{1}+$ $\ldots+c_{n}(t) \Psi_{n}$, where $\Psi_{h}$ is an orthonormal basis in $H_{1}$ and the system of equations of the Bubnov-Galerkin method is solvable at each step and the sequence $\mathbf{u}_{n}{ }^{*}$ converges strongly to the solution of the problem in the space $B_{1}^{n, n}(T)$.

By estimating the members in the right side of (3.2) by using the Hölder inequality, it is easy to see that each of these members is a linear continuous functional in the variable $\delta \mathbf{u}=\varphi$ in the space $H_{2}(T)$. By the Riesz theorem on the representation of a continuous linear functional in Hilbert space, (3.2) can be written in the equivalent operator form

$$
\begin{equation*}
(\mathbf{v} \cdot \varphi)_{H_{2}(T)}=\left(G_{T} \mathbf{V} \cdot \varphi\right)_{H_{2}(T)}+(\mathbf{f} \cdot \varphi)_{H_{2}(T)} \tag{3.3}
\end{equation*}
$$

The following iteration process is constructed

$$
\mathbf{v}_{n}=G_{T} \mathbf{v}_{n-1}+\mathbf{f}
$$

where $\mathbf{v}_{0}$ is an arbitrary vector function, $\mathbf{v}_{0} \equiv H_{2}(T)$. We call the corresponding process for (1.3) the method of "elastic solutions". By using the Hölder inequality, the characteristic term in the expression $\left(G_{T} \vee \cdot \varphi\right)_{H_{2}(T)}$ is estimated as follows:

$$
\begin{aligned}
& \left|\int_{0}^{T} \int_{\Omega}^{\tau} \int_{0}^{\tau} E(\tau-\alpha) \varepsilon(\alpha) d \alpha \varepsilon_{1}(\tau) d \Omega d \tau\right| \leqslant \\
& m_{1}\left\{\int_{0}^{T} t \int_{\Omega}^{t} \int_{0}^{t} \varepsilon^{2}(x) d \alpha d \Omega d t\right\}^{1 / 2}\left\|\varepsilon_{1}\right\|_{L_{2}(\Omega \times[0, T])}
\end{aligned}
$$

where

$$
m_{1}=\max |E(t)| \text { for } t \in[0, T]
$$

It must be noted that $T$ is arbitrary in this inequality. Summing the estimates for all the members of $\left(G_{T} \mathbf{V} \cdot \varphi\right)_{H_{2}(T)}$, the estimate

$$
\begin{equation*}
\left\|G_{l} \mathbf{v}\right\|_{H_{2}(t)} \leqslant m_{2}\left\{\int_{0}^{t} \tau\|\mathbf{v}\|_{H_{2}(\tau)}^{2} d \tau\right\}^{2 / 2} \tag{3.4}
\end{equation*}
$$

can be obtained by selecting $\varphi=G_{T} \mathbf{v}$ and using Lemma 2.1 , where the constant $m_{2}$ depends only on $t$ and is finite for $0 \leqslant t \leqslant T<\infty$. The estimate

$$
\begin{equation*}
\left\|G_{T}^{k+1} \mathbf{v}\right\|_{I_{z}(T)} \leqslant\|\mathbf{v}\|_{H_{z}(T)}\left(2^{-1 / 2} m_{2} T\right)^{k}(k!)^{-1} \tag{3.5}
\end{equation*}
$$

is deduced from (3.4), from which there results that for sufficiently large $k$ the operator $G T^{k}$ is a compression operator. According to the principle of compressed mappings (3.3) is uniquely solvable. By using the estimate (3.5) the rate of convergence of the series $v_{0}+\left(v_{1}-v_{0}\right)+\left(v_{2}-v_{1}\right)+\ldots$ to (3.3) is found. Therefore, the following theorem is valid, from which the first part of the assertion in Theorem 3.1 follows.

Theorem 3.2. Let all the conditions of Theorem 3.1 be satisfied. In this case the sequence of approximate solutions of the method of elastic solutions reduces to a single generalized solution of the linear viscoelasticity problem $u$ with the velocity

$$
\left\|\mathbf{u}-\mathbf{u}_{m}\right\|_{\mathbf{B}_{1}^{n, n}(T)} \leqslant K \sum_{k=m}^{\infty} \frac{(M T)^{k}}{k!}
$$

Here the constant $M$ depends only on $T$, and the constant $K$ on the magnitudes of the corresponding norms of the external forces, the vector function $\boldsymbol{\Phi}$, the initial approximation $\mathbf{u}_{0} \in B_{1}^{n, n}(T)$, constants from the imbedding theorem; $M$ and $K$ are independent of $m$.

The system of $r$ th approximation equations of the Bubnov-Galerkin method in the terminology of (3.2) is

$$
\left(\mathbf{v}_{r}^{*} \cdot \boldsymbol{\Psi}_{k}\right)_{H_{1}}=R_{1}\left(\mathbf{v}_{r}^{*}, \boldsymbol{\Psi}_{k}\right)+R_{2}\left(\Phi, \mathbf{F}, \mathbf{P}, \boldsymbol{\psi}_{k}\right), \quad k=1, \ldots, r
$$

The solvability of this integro-differential system for any $r$ is proved just as the solvability of (3.2). The estimate of the norm of the solution $\mathbf{v}_{r} *$ obtained is hence independent of the number $r$. Hence, a weakly convergent subsequence in $H_{2}(T)$ can be selected from the sequence $\mathbf{v}_{r}{ }^{*}$. It is easy to show that this weak limit is indeed the generalized solution of the problem. By virtue of the uniqueness of the solution of the problem the whole sequence converges weakly. Strong convergence of the sequence $\mathbf{v}_{r}{ }^{*}$ in the small time segment $\left[0, T_{1}\right]$ follows from the fact that $G_{t}$ is a compression operator for sufficiently small $t$. Now, examining the segment $\left[T_{1}, 2 T_{1}\right]$ and taking account of the properties of the operator $G_{T}$ and the strong convergence of $\mathbf{v}_{r}{ }^{*}$ in the space $H_{2}\left(T_{1}\right)$, it can be seen that the sequence $\mathbf{v}_{T}{ }^{*}$ converges strongly in the norm $H_{2}\left(2 T_{1}\right)$, etc., until the whole segment $[0, T]$ is exhausted in a finite number of steps.

Note 1. Let $\boldsymbol{\Phi} \in C^{n}\left[0, T ; H_{1}\right]$ and in the variable $\varphi$

$$
\int_{S} \varphi_{i} C\left(\partial_{t}\right) P_{i} d S+\int_{\Omega} \varphi_{i} C\left(\partial_{i}\right) F_{i} d \Omega
$$

is a continuous operator in the space $C^{\circ}\left[0, T ; H_{1}\right]$ in the space of continuous functions. In this case the generalized solution of the problem (from Theorem 3) lies in the space $C^{n}\left[0, T ; H_{1}\right]$. This note results from Theorem 1 formulated in [4].

Note 2. In the case of the so-called quasi-static problem (obtained formally from (1.3) for $\rho=0$ ) Theorems 3.1,3.2 are true, where compliance with the condition $m_{C}+2 \leqslant m_{B}$ is optional.
4. Let the vector function be $\varphi \in H_{3}{ }^{\circ}(T)$, and for definiteness $m_{A}=m_{B}$. Then (1.3) is converted by the following two methods:

$$
\begin{gather*}
\left(\partial_{i}^{n} \mathbf{u} \cdot \varphi\right)_{H_{2}(T)}-\int_{0}^{T}\left\{\left(\rho \partial_{t}^{n+1} \mathbf{u} \cdot \partial_{i} \varphi\right)_{L_{2}(\Omega)}-\left(\rho c_{1} \partial_{t}^{n+1} \mathbf{u} \cdot \varphi\right)_{L_{2}(\Omega)}-\right.  \tag{4.1}\\
\left.\left(\rho c_{2} \partial_{t}^{n} \mathbf{u} \cdot \varphi\right)_{L_{2}(\Omega)}\right\} d t=-\left.\left(\rho \mathbf{a}_{n+1} \cdot \varphi\right)_{L_{2}(\Omega)}\right|_{t=0}+\int_{0}^{T} R_{3}(\mathbf{u}, \mathbf{F}, \mathbf{P}, \varphi) d t \\
\text { for } n=m_{B}=m_{C} \\
\left(\partial_{l}^{n} \mathbf{u} \cdot \varphi\right)_{H_{2}(T)}-\int_{0}^{T}\left\{\left(\rho \partial_{t}^{n} \mathbf{u} \cdot \partial_{t} \varphi\right)_{L_{2}(\Omega)}-\left(\rho c_{1} \partial_{t}^{n} \mathbf{u} \cdot \varphi\right)_{L_{2}(\Omega)}\right\} d t=  \tag{4.2}\\
\left.\left(\rho \mathbf{a}_{n} \cdot \varphi\right)_{L_{2}(\Omega)}\right|_{t=0}+\int_{0}^{T} R_{\mathbf{4}}(\mathbf{u}, \mathbf{F}, \mathbf{P}, \varphi) d t \quad \text { for } n=m_{B}=m_{C}+1
\end{gather*}
$$

Here $R_{3}, R_{4}$ are linear forms in each of their variables containing $\partial_{t}{ }^{k} \mathbf{u}$ only for $k<n$.

Definition 4.1. The vector function $\mathbf{u} \in B_{1}^{n, n+1}(T)\left[\mathbf{u} \in B_{3}{ }^{n}(T)\right]$ satisfying the initial and boundary conditions (1.4), (1.5) and the equalities (4.1), (4.2) for any vector function $\varphi \in H_{3}{ }^{\circ}(T)$, where the $n$th initial condition is taken in the following sense

$$
\lim \left\|\partial_{t}^{n} \mathbf{u}-\mathbf{a}_{n}\right\|_{L_{t}(\Omega)}=0 \quad \text { for } \quad t \rightarrow 0
$$

is called the generalized solution of the linear viscoelasticity problem in the case $n=m_{B}=m_{C}\left[n=m_{B}=m_{C}+1\right]$.

Theorem 4.1. Let the following conditions be satisfied:

1) $n=m_{B} \geqslant m_{A}$,
2) $\mu>0, \rho>0$,
3) $\lambda+2 \mu>0$, if $m_{B}=m_{A}$,
4) $\boldsymbol{\Phi} \in B_{1}^{n+l, n+l+1}(T)$, if $n=m_{C}$, 5) $\boldsymbol{\Phi} \in B_{1}^{n+l, n+l} \quad(T)$, if $n=m_{C}+1$,
5) $\left.\partial_{t}^{k-1} C\left(\partial_{t}\right) F_{i} \in L_{2}(\Omega \times[0, T]), \quad 7\right) \partial_{t}{ }^{k} C\left(\partial_{t}\right) P_{i} \in L_{2}(S \times[0, T])$,

$$
\text { 8) } k=0, \ldots, l, \quad l \geqslant 1
$$

In the case $n=m_{C}\left[n=m_{C}+1\right]: 1$ ) there exists a single generalized solution of the linear viscoelasticity problem in the sense of the definition $4.1 ; 2$ ) the solution can be found by the Bubnov-Galerkin method, where the system of equations of the Bub-nov-Galerkin method is uniquely solvable at each step, the sequence of approximate solutions of the Bubnov-Galerkin method lies in some sphere of the space $B_{1}^{n+l-1, n+l}(T)$ $\left[B_{3}^{n+l-1}(T)\right]$ and converges weakly to the generalized solution of the problem $\mathbf{u} \in$ $B_{1}^{n+l-1, n+l}(T)\left[\mathbf{u} \in B_{3}^{n+l-1}(T)\right]$; 3) if $l \geqslant 2$, the sequence $\mathbf{u}_{r}$ then converges strongly in the space $B_{1}^{n+l-2, n+l-1}(T)\left[B_{3}^{n+l-2}(T)\right]$.

The proof of the theorem of the existence of the Bubnov-Galerkin method for nonstationary problems was first proposed in [5].

The course of the proof in the case $n=m_{C}$ is described briefly below. As in Sect. 3 , the solution is sought in the form

$$
\begin{aligned}
& \text { the form } \\
& \mathbf{u}=\mathbf{u}_{0}+\boldsymbol{\Phi}, \quad \mathbf{u}_{0}=\int_{0}^{t} \frac{(t-\tau)^{n-1}}{(n-1)!} \mathbf{v} d \tau
\end{aligned}
$$

which reduces to an equivalent problem whose unique solvability involves the existence of a single generalized solution of the linear viscoelasticity problem.

Equivalent problem. Find the vector function $v \in H_{3}(T)$ satisfying the following equality

$$
\begin{align*}
& \quad(\mathbf{v} \cdot \varphi)_{H_{2}(T)}-\int_{0}^{T}\left\{\left(\rho \partial_{t} \mathbf{v} \cdot \partial_{t} \varphi\right)_{L_{2}(\Omega)}-\left(\rho c_{1} \partial_{t} \mathbf{v} \cdot \varphi\right)_{L_{2}(\Omega)}-\right. \\
& \left.\left(\rho C_{2} \mathbf{v} \cdot \varphi\right)_{L_{2}(\Omega)}\right\} d t=\int_{0}^{T}\left\{R_{3}\left(\int_{0}^{t} \frac{(t-\tau)^{n-1}}{(n-1)!} \mathbf{v} d \tau, \mathbf{F}, \mathbf{P}, \varphi\right)+R_{5}(\Phi, \varphi)\right\} d t \tag{4.3}
\end{align*}
$$

for any vector function $\varphi \in H_{3}^{0}(T)$ and $\mathbf{v}=0$ for $t=0$.
The proof of the existence of the solution of the equivalent problem is by the BubnovGalerkin method; the $r$ th approximation of the method

$$
\mathbf{v}_{r}=\sum_{k=1}^{r} a_{r k}(t) \psi_{k}
$$

is found from the system of equations

$$
\begin{align*}
& \alpha_{r k}^{\prime \prime}(t)+c_{1} \alpha_{r k}^{\prime}(t)+c_{2} \alpha_{r k}(t)+\left(\mathbf{v}_{r} \cdot \boldsymbol{\psi}_{h}\right)_{H_{1}}= \\
& R_{3}\left(\int_{0}^{t} \frac{(t-\tau)^{n-1}}{(n-1)!} \mathbf{v}_{r} d \tau, \mathbf{F}, \mathbf{P}, \boldsymbol{\psi}_{k}\right)+R_{5}\left(\boldsymbol{\Phi}, \boldsymbol{\varphi}_{k}\right) \tag{4.4}
\end{align*}
$$

with the initial conditions $\alpha_{r k}(0)=\alpha_{r k}^{\prime}(0)=0, k=1, \ldots, r$. Here the $\psi_{k}$ are elements of the basis of the space $H_{1}$ such that

$$
\mathrm{P}\left(\boldsymbol{\Psi}_{k} \cdot \boldsymbol{\varphi}_{n}\right)_{L_{z}(\Omega)}=\delta_{k n}
$$

The uniqueness of solution of the system of Bubnov-Galerkin Eqs. (4.4) follows from the
fact that a system of ordinary differential equations of normal type with constant coefficients is obtained for the corresponding coefficients $\alpha_{r_{k}}$ of the vector function

$$
\mathbf{u}_{0 r}=\int_{0}^{t} \frac{(t-\tau)^{n-1}}{(n-1)!} \mathbf{v}_{r} d \tau
$$

The a priori estimate of the solution (4.4) is proved in the space $H_{3}(T)$, for which the $k$ th equation of (4.4) is multiplied by $\alpha_{r k^{\prime}}$ and these equalities are then added and integrated with respect to the time between the limits 0 and $t$. A vector function is contained in the right sides of (4.4) only under the time integral, hence integrating by parts with respect to $t$ in the equality obtained, we have

$$
\left|\int_{0}^{t} R_{3}\left(\mathbf{u}_{0 r}, \mathbf{F}, \mathbf{P}, \partial_{\tau} \mathbf{v}_{r}\right) d \tau\right| \leqslant M_{1} \int_{0}^{t}\left\|\mathbf{v}_{r}\right\|_{H_{1}}^{2} d \tau+M_{2}\left\|\partial_{\tau} \mathbf{v}_{r}\right\|_{L_{2}(\Omega \times[0, T])}+0.5\left\|\mathbf{v}_{r}\right\|_{H_{1}}^{2}
$$

The left side of the equality obtained is of the form

$$
\left\|\partial_{t} \mathbf{v}_{r}\right\|_{L_{2}(\Omega)}^{2}+c_{1} \int_{0}^{t}\left\|\partial_{\tau_{r}} \mathbf{v}_{r}\right\|_{L_{2}(\Omega)}^{2} d \tau+c_{2} \int_{0}^{t}\left(\mathbf{v}_{r} \cdot \partial_{\tau} \mathbf{v}_{r}\right)_{L_{2}(\Omega)} d \tau+\left\|\mathbf{v}_{r}\right\|_{H_{1}}^{2}
$$

A uniform estimate for all $0 \leqslant t \leqslant T$, independent of $r$

$$
\begin{equation*}
\left\|\mathbf{v}_{r}\right\|_{H_{1}}^{2}+\left\|\partial_{t} \mathbf{v}_{r}\right\|_{L_{2}(\Omega)}^{2} \leqslant M_{3} \tag{4.5}
\end{equation*}
$$

results from a comparison between the right and the left sides of the equality and the Gronwall inequality. Here the constants $M_{k}, k=1,2,3$ depend only on the magnitudes of the corresponding norms $\boldsymbol{\Phi}, \mathbf{P}, \mathbf{F}$ and the length of the segment $[0, T]$. By virtue of the a priori estimate (4.5), a subsequence converging weakly to $v_{0}$ in the space $H_{3}(T)$ can be selected from the sequence of approximations $\mathbf{v}_{r}$. The imbedding operator $H_{3}(T)$ in the space $L_{2}(\Omega)$ is completely continuous, hence $\mathrm{v}_{0}=0$ for $t=0$.

The system (4.4) reduces to the form (4.3) for which the $k$ th equality in (4.4) is multiplied by $d_{k}(t)$, the equalities are then added, integrated with respect to time, and where necessary integrated by parts with respect to time. A passage to the limit shows that $\mathbf{v}_{0}$ satisfies (4.3) for any function $\varphi$ of the form

$$
\begin{equation*}
\varphi=\sum_{k=0}^{N} d_{k}(t) \psi_{k}, \quad d_{k}(T)=0 \tag{4.6}
\end{equation*}
$$

Because the set of functions of the form (4.6) is compact in $H_{3}{ }^{\circ}(T)$, there results that $\mathbf{v}_{0}$ is a solution of the equivalent problem. To prove the uniqueness of the solution, we set

$$
\varphi=\chi(t)=\int_{t}^{8} v(\tau) d \tau, \quad \chi(t)=0, \quad \text { if } \quad t \geqslant s
$$

in (4.3) for $\mathbf{F}=0, \mathbf{P}=0, \Phi=0$. After elementary estimates similar to those made to obtain the inequality (4.5), we obtain

$$
\|\chi(0)\|_{H_{1}}^{2}+\|\mathbf{v}(s)\|_{L_{2}(\Omega)}^{2} \leqslant M_{4}\left\{\int_{0}^{8}\|\chi(t)\|_{H_{3}(t)}^{2} d t+\|\chi(0)\|_{L_{2}(\Omega)}^{2}\right\}
$$

from which the uniqueness of the solution of the equivalent problem results. It follows from the uniqueness of the solution that the whole sequence of approximations $\mathbf{v}_{r}$ of the Bubnov-Galerkin method converges weakly in $H_{3}(T)$.

The reasoning relative to Eq. (4.4) differentiated $l-1$ times with respect to time is analogous to that presented above, and terminates the proof of the second part of Theorem 4.1. The strong convergence of the subsequence of approximate solutions of the Bubnov-Galerkin method can be shown by proceeding from the density of a function of the form

$$
\sum_{k=0}^{N} c_{k}(t) \Psi_{k}, \quad c_{k}(t) \in C^{\infty}(0, T)
$$

in the space $B_{1}^{n, m}(T) \cap B_{2}^{n}(T)$. The case $n=\bar{m}_{C}+1$ is examined analogously : the problem is reduced to the equivalent problem by the substitution (3.1). A certain difference in the investigation of the equivalent problem from that carried out above for $n=m_{\mathcal{C}}$ is the following: firstly, to obtain the a priori estimate of solutions of the Bubnov-Galerkin system it is necessary to multiply the $k$ th equation of the system by $\alpha_{r k}$, but not by $\alpha_{r k^{\prime}}$, and secondly, it is necessary to prove boundedness in the norm $L_{2}(\Omega)$ for all $0 \leqslant t \leqslant T$ for the limit element of the Bubnov-Galerkin sequence. The method of investigating the equivalent problem agrees completely with the corresponding method of investigating a linear parabolic equation [6].

Note 1. Conditions (6), (7) of Theorem 1 can be replaced by the following

$$
\begin{gathered}
\partial_{t}^{k} C\left(\partial_{t}\right) \mathbf{F} \in L_{2}\left[0, T ; H_{1}^{-}\right] \\
\partial_{t}^{k} C\left(\partial_{t}\right) P_{i} \in L_{2}\left[0, T ; W_{2}^{-1 / 2}(S)\right]
\end{gathered}
$$

Note 2. All the results obtained above are valid for the plane viscoelasticity problem.

Note 3. All the results obtained above can be extended to the case of an anisotropic body, where the coefficients of the appropriate polynomials can depend piecewisecontinuously on the coordinates $x_{i}$ and time $t$.

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