# AN IDENTIFICATION PROBLEM ARISING IN THE THEORY OF HEAT CONDUCTION FOR MATERIALS WITH MEMORY 

A. FAVARON<br>Dipartimento di Matematica " $F$. Enriques", Universitá di Milano, Via Saldini 50, 20133 Milano, Italy E-mail: favaron@mat.unimi.it


#### Abstract

We deal with the problem of recovering a memory kernel $k(t, \eta)$, depending on time $t$ and on a spatial variable $\eta$, in a parabolic integro-differential equation related to a bounded domain $\Omega \subset \mathbf{R}^{3}$. We show that, under suitable assumptions and two pieces of additional information, our identification problem can be uniquely solved locally in time.


## 1. INTRODUCTION

We deal with an identification problem arising when the classical theory of heat conduction is modified in order to describe the thermal behaviour of viscoelastic materials such as polymers, polymer solutions and suspensions (cf. [9]). With respect to metals, the mechanical properties of viscoelastic materials can be affected by the previous history, that is, by the method of fabrication, post-treatment, and the age of the finished article. This is why these materials are said to possess memory. Such a memory evinces in the constitutive relationship between the stress and the strain tensor and leads to mathematical models of viscoelastic phenomena which take the form of partial differential Volterra equations (cf. [1], [6] and [10]).

Indeed, if $\Omega \subset \mathbf{R}^{3}$ is a nonhomogeneous thermal body made of material with memory, then the variation of the temperature $u$ with respect to time satisfies the following parabolic integro-differential equation, where the symbol "*" stands for the convolution operator $(v * w)(t)=\int_{0}^{t} v(t-s) w(s) \mathrm{d} s$ :

$$
\begin{equation*}
D_{t} u(t, x)=\mathcal{A} u(t, x)+\operatorname{div}[(k(\cdot, \rho(x)) * \widetilde{b}(x) \nabla u(\cdot, x))(t)] \mathrm{d} s+f(t, x), \quad t>0, x \in \Omega \tag{1}
\end{equation*}
$$

Here $\mathcal{A}$ is a second-order linear differential operator, $k$ represents the memory kernel keeping the history record of the material, $\rho: \Omega \rightarrow \mathbf{R}$ is an assigned function, $f$ represents the system of external heat sources, $\widetilde{b}(x)$ denotes a $3 \times 3$ matrix $\left(\widetilde{b}_{i, j}(x)\right)_{i, j=1}^{3}$ and, finally, div $=\sum_{i=1}^{3} D_{x_{i}}$ and $\nabla=\left(D_{x_{1}}, D_{x_{2}}, D_{x_{3}}\right)$.

In this paper, we will deal with the problem of recovering $k$ in equations of the type of (1). It should be stressed that the recovery of a memory kernel $k$ depending on both time and space is a quite new problem, as far as first-order in time integro-differential equations are concerned. See, for instance, [2], [5] and [7] which, however, have to be considered one-dimensional in character. In [2] the kernel $k$ depends on time and on only one space variable between the $n$ variables of $\mathbf{R}^{n}, n \geq 2$, whereas in [5] and [7] the kernel is assumed to be degenerate, i.e. of the form $k(t, x)=\sum_{j=1}^{N} m_{j}(t) \mu_{j}(x)$, but with the spacedependent functions $\mu_{j}, j=1, \ldots, N$, assumed to be known, too. As a consequence, the identification problem reduces to recovering the $N$ unknown time-dependent functions $m_{j}, j=1, \ldots, N$. Our aims will be those of generalizing these one-dimensional results, by skipping the assumption of degenerateness and searching for kernels which depend on both time and space, the spatial dependence occurring through scalar functions of all the variables at disposal.

We point out that the investigation of time and space dependent memory kernels seems to be very promising from the point of view of applications. For example, it gives a quite good description of the thermal behaviour of nonhomogeneous bodies and highlights the existence of unexpected symmetry relationships, between the analytic properties of kernels and the geometric structure of materials (cf. [3]).

## 2. FORMULATION OF THE IDENTIFICATION PROBLEM

Our problem is concerned with the identification of the unknown memory kernel $k$, depending on two scalar variable $t, \eta$, appearing in the following integro-differential equation of parabolic type related to a bounded domain $\Omega \subset \mathbf{R}^{3}$, where $(t, x) \in[0, T] \times \Omega$ :

$$
\begin{equation*}
D_{t} u(t, x)=\mathcal{A} u(t, x)+[k(\cdot, \rho(x)) * \mathcal{B} u(\cdot, x)](t)+\left[D_{\eta} k(\cdot, \rho(x)) * \mathcal{C} u(\cdot, x)\right](t)+f(t, x) . \tag{2}
\end{equation*}
$$

Here, $\mathcal{A}$ and $\mathcal{B}$ are two second-order linear differential operators, while $\mathcal{C}$ is a first-order differential
operator, having respectively the following forms:

$$
\begin{equation*}
\mathcal{A}=\sum_{j, k=1}^{3} D_{x_{j}}\left(a_{j, k}(x) D_{x_{k}}\right), \quad \mathcal{B}=\sum_{j, k=1}^{3} D_{x_{j}}\left(b_{j, k}(x) D_{x_{k}}\right), \quad \mathcal{C}=\sum_{j=1}^{3} c_{j}(x) D_{x_{j}} \tag{3}
\end{equation*}
$$

In addition, the operator $\mathcal{A}$ is uniformly elliptic whereas the function $\rho$ satisfies the following assumptions:
(A) $\rho: V \rightarrow \rho(V) \subset \mathbf{R}$, where $V$ is an open (possibly unbounded) set of $\mathbf{R}^{3}$ and $\rho(V)=\left(l_{1}, l_{2}\right)$, $-\infty \leq l_{1}<l_{2} \leq+\infty ;$
(B) if $\bar{V}=V \cup \partial V$ denotes the $\mathbf{R}^{3}$-closure of $V$, then $\partial V=\bigcup_{i=1}^{2} \bar{U}_{i}$ where $U_{i}=\left\{x \in \mathbf{R}^{3}\right.$ : $\left.(-1)^{i} \limsup _{y \rightarrow x, y \in V}(-1)^{i} \rho(y)=l_{i}\right\}, i=1,2 ;$
(C) $\rho \in C^{1}(V)$ and $\nabla \rho(x) \neq 0$ for every $x \in V$;
(D) for any $y \in \partial V$ there exist $r_{y}>0$ and at least an index $j(y) \in\{1,2,3\}$ such that $D_{x_{j(y)}} \rho(z) \neq 0$ for every $z \in V \cap\left(B\left(y, r_{y}\right) \backslash U_{y}^{j(y)}\right)$, where $U_{y}^{j(y)}$ is a subset of $\mathbf{R}^{3}$ having three-dimensional Lebesgue measure equal to zero.

Observe that, in (B), one or both of the $\bar{U}_{i}$ 's could be empty (for instance if $V=\mathbf{R}^{3}$ ) or they could coincide (for instance if $\partial V$ is an hyperplane and $V$ stands on both the sides of $\partial V$ ). Moreover, in (D), when $U_{y}^{j(y)}$ is contained in $\partial V$ we can choose $U_{y}^{j(y)}=\emptyset$.

However, it is worth saying that assumptions (A)-(C), except for $\nabla \rho(\cdot) \neq 0$ in $V$ and at least when $\bar{U}_{1} \cap \bar{U}_{2}=\emptyset$, are close to requiring $\rho \in C(\bar{V}) \cap C^{1}(V)$ and hence are quite natural. On the contrary, assumption (D) seems a very special one and hard to be satisfied. This is not completely true and, as a concrete example, we refer to [3, Examples 3.1.2] where it is shown that (A)-(D) are all satisfied by each one of the spherical coordinates $(r, \varphi, \theta)$, related to the Cartesian ones via the relationship $\left(x_{1}, x_{2}, x_{3}\right)=r(\cos \varphi \sin \theta, \sin \varphi \sin \theta, \cos \theta)$.

Coming back to our problem, we restrict our attention to any bounded domain $\Omega$ of $\mathbf{R}^{3}$ such that $\bar{\Omega} \subset \bar{V}$ and such that the boundary $\partial \Omega$ is the finite union of $m$ pairwise disjoint surfaces $\partial \Omega_{k}, k=1, \ldots, m$, of class $C^{2}$, i.e. $\partial \Omega=\bigcup_{k=1}^{m} \partial \Omega_{k}, \partial \Omega_{i} \cap \partial \Omega_{j}=\emptyset$ for any $i, j=1, \ldots, m, i \neq j$. However, a bounded domain $\Omega$ of $\mathbf{R}^{3}$ satisfying the previous assumptions will be said an admissible domain for our problem if and only if the function $\rho$ satisfies the following additional property on it:
(E) there exists a constant $\widetilde{C}$ such that $|\nabla \rho(x)| \leq \widetilde{C}$ for every $x \in \bar{\Omega} \cap V$.

Note that (E) is obvious if $\bar{\Omega} \subset V$, much less if $\partial \Omega \cap \partial V \neq \emptyset$. Now, $u_{0}: \bar{\Omega} \rightarrow \mathbf{R}$ and $u_{1}:[0, T] \times \bar{\Omega} \rightarrow \mathbf{R}$ being two assigned smooth functions, we prescribe the initial condition

$$
\begin{equation*}
u(0, x)=u_{0}(x), \quad \forall x \in \Omega \tag{4}
\end{equation*}
$$

as well as one of the following boundary conditions, where, for any $s \in\{1, \ldots, m\}$, the indexes $i_{k}$, $k=1, \ldots, s$, satisfy $0 \leq i_{1}<\ldots<i_{s} \leq m$ with the convention that $\partial \Omega_{i_{1}}=\emptyset$ when $i_{1}=0$ :

$$
\begin{array}{ll}
\left(\mathrm{D} ; i_{1}, \ldots, i_{s}\right) & u(t, x)=u_{1}(t, x), \quad \forall(t, x) \in[0, T] \times \bigcup_{k=1}^{s} \partial \Omega_{i_{k}} \\
\left(\mathrm{~N} ; i_{1}, \ldots, i_{s}\right) & D_{\nu} u(t, x)=D_{\nu} u_{1}(t, x), \quad \forall(t, x) \in[0, T] \times\left(\partial \Omega \backslash \bigcup_{k=1}^{s} \partial \Omega_{i_{k}}\right) \tag{6}
\end{array}
$$

Here $\nu$ stands for the conormal vector $\nu=\left(\nu_{1}, \nu_{2}, \nu_{3}\right)$ being defined by $\nu_{j}(x)=\sum_{k=1}^{3} a_{j, k}(x) n_{k}(x)$, where $n(x)=\left(n_{1}(x), n_{2}(x), n_{3}(x)\right)$ denotes the unit outer normal vector at $x \in \partial \Omega$.

Having to deal with the inverse problem of determining both $u$ and $k$ in (2), it seems to be reasonable to prescribe one more additional information, to be used for reconstructing $k$, and to hope that only one may be enough. By the way, this is not the case, since the kernel $k$ we want to recover explicitly depends on the variables $t$ and $\eta=\rho(x), x \in \Omega \cap V$, and therefore the problem is two-dimensional in character. Indeed, by setting

$$
\left\{\begin{array}{l}
l_{3}=\inf _{x \in \Omega \cap \bar{V}}\left\{\liminf _{y \rightarrow x, y \in \Omega \cap V} \rho(y)\right\}  \tag{7}\\
l_{4}=\sup _{x \in \Omega \cap \bar{V}}\left\{\limsup _{y \rightarrow x, y \in \Omega \cap V} \rho(y)\right\}
\end{array}\right.
$$

fixing $\eta_{0} \in\left[l_{3}, l_{4}\right]$ and assuming $k(t, \eta)$ differentiable with respect to $\eta$, from the fundamental formula of Integral Calculus we get

$$
\begin{equation*}
k(t, \eta)=h(t)+\int_{\eta_{0}}^{\eta} q(t, \xi) \mathrm{d} \xi=: h(t)+E q(t, \eta), \quad \forall(t, \eta) \in[0, T] \times\left(l_{3}, l_{4}\right) \tag{8}
\end{equation*}
$$

where $h$ and $q$ denote the new unknown functions

$$
\begin{equation*}
h(t)=k\left(t, \eta_{0}\right), \quad q(t, \eta)=D_{\eta} k(t, \eta) \tag{9}
\end{equation*}
$$

Consequently, recovering $k$ is equivalent to recovering $h$ and $q$ and, as we will see further, the original problem of identifying the pair $(u, k)$ in equation (2) can be reformulated as an equivalent one related to the triplet $(v, h, q)$ where $v=D_{t} u-D_{t} u_{1}$. Hence, it clearly appears that one additional piece of information does not suffice to solve our identification problem, and, by virtue of decomposition (8), at least two additional measurements have to be available. Therefore, we assume also to be given the two following additional pieces of information

$$
\begin{align*}
& \Phi[u(t, \cdot)](\eta):=g_{1}(t, \eta), \quad \forall(t, \eta) \in[0, T] \times\left(l_{3}, l_{4}\right),  \tag{10}\\
& \Psi[u(t, \cdot)]:=g_{2}(t), \quad \forall t \in[0, T] \tag{11}
\end{align*}
$$

where $\Phi, \Psi$ and $g_{i}, i=1,2$, are, respectively, given linear operators acting on spatial variables and smooth prescribed functions.

By writing $\omega=\omega(\cdot)$ when this shortening does not generate any confusion, from (4)-(6), (10), (11) we (formally) deduce that our data have to satisfy also the following consistency conditions

$$
\begin{array}{ll}
(\mathrm{C} 1, \mathrm{D}, \mathrm{~N}) \quad u_{0}=u_{1}(0, \cdot) \quad \text { on } \Gamma_{\mathrm{D}} \quad \text { and } \quad D_{\nu} u_{0}=D_{\nu} u_{1}(0, \cdot) \quad \text { on } \Gamma_{\mathrm{N}} \\
& \Phi\left[u_{0}\right](\eta)=g_{1}(0, \eta), \quad \Psi\left[u_{0}\right]=g_{2}(0) \tag{13}
\end{array}
$$

$\Gamma_{\mathrm{D}}$ and $\Gamma_{\mathrm{N}}$ in (12) denoting, respectively, those parts of the boundary of $\Omega$ where Dirichlet and Neumann conditions are possibly prescribed.

Convention: From now on, when $\Omega$ is an admissible domain, we will denote by $\mathrm{P}(\mathrm{D}, \mathrm{N})$ the identification problem consisting of (2), (4), the boundary conditions (5), (6) and the additional pieces of information (10), (11).

## 3. MAIN ASSUMPTIONS AND BASIC SYSTEM

Since we are interested in an existence and uniqueness result to the identification problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$, the working technique seems to be that of reformulating our problem in a Banach space framework and then to apply a fixed-point argument to such a problem (cf. [2]).

To perform this procedure, we first rewrite our problem for the pair $(u, k)$ in an equivalent one for the triplet $(v, h, q)$, where $h$ and $q$ are defined by (9) whereas $v$ is defined via the following formula

$$
\begin{equation*}
v:=D_{t} u-D_{t} u_{1} \quad \Longleftrightarrow u(t, x)=u_{1}(t, x)-u_{1}(0, x)+u_{0}(x)+\int_{0}^{t} v(s, x) \mathrm{d} s \tag{14}
\end{equation*}
$$

$u_{0}$ and $u_{1}$ being, respectively, the prescribed initial and boundary data (4)-(6).
To obtain the basic system for the triplet $(v, h, q)$ we need some further assumptions, additional to the linearity, for the operators $\Phi$ and $\Psi$ appearing in (10) and (11). We now list the essential ones. Let $\Xi$ be the set of spatial variables on which $\Phi$ acts and let $v_{1}:\left(l_{3}, l_{4}\right) \rightarrow \mathbf{R}, v_{2}: \Omega \rightarrow \mathbf{R}$ and $v_{3}:\left(l_{3}, l_{4}\right) \times \Xi \rightarrow \mathbf{R}$ be three functions for which at the moment we do not require any other regularity but that $v_{3}$ admits (generalized) derivative with respect to $\eta \in\left(l_{3}, l_{4}\right)$. Then the following relations hold:

$$
\begin{equation*}
\Phi\left[v_{1} v_{2}\right]=v_{1} \Phi\left[v_{2}\right], \quad D_{\eta} \Phi\left[v_{3}\right](\eta)=\Phi\left[D_{\eta} v_{3}\right](\eta), \quad \forall \eta \in\left(l_{3}, l_{4}\right) \tag{15}
\end{equation*}
$$

It occurs that (15) is not sufficient and we need also the two basic decomposition formulae

$$
\begin{equation*}
\Phi \mathcal{A}=\mathcal{A}_{1} \Phi+\Phi_{1}, \quad \Psi \mathcal{A}=\Psi_{1} \tag{16}
\end{equation*}
$$

where $\Phi_{1}$ and $\Psi_{1}$ are two linear operators and $\mathcal{A}_{1}$ is a differential operator $\mathcal{A}_{1}\left(\eta ; D_{\eta}\right)$.

Further, we assume that $u_{0}$ satisfies the following conditions for some positive constant $m_{0}$ :

$$
\begin{align*}
J_{0}\left(u_{0}\right)(\eta) & :=\left|\Phi\left[\mathcal{C} u_{0}\right](\eta)\right| \geq m_{0}, \quad \forall \eta \in\left(l_{3}, l_{4}\right)  \tag{17}\\
J_{1}\left(u_{0}\right) & :=\Psi\left[J\left(u_{0}\right)\right] \neq 0 \tag{18}
\end{align*}
$$

where we have set:

$$
J\left(u_{0}\right)(x):=\left(\mathcal{B} u_{0}(x)-\frac{\Phi\left[\mathcal{B} u_{0}\right](\rho(x))}{\Phi\left[\mathcal{C} u_{0}\right](\rho(x))} \mathcal{C} u_{0}(x)\right) \exp \left[\int_{\rho(x)}^{\eta_{0}} \frac{\Phi\left[\mathcal{B} u_{0}\right](\xi)}{\Phi\left[\mathcal{C} u_{0}\right](\xi)} \mathrm{d} \xi\right], \quad \forall x \in \Omega
$$

Now, let us suppose that the pair $(u, k)$ is a solution to the identification problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$ such that $u$ is twice differentiable with respect to time and $k$ is once differentiable with respect to $\eta$. Replacing $k$ in (2) with the right-hand side of (8) and differentiating the so obtained equation with respect to time, we deduce that the triplet $(v, h, q)$ satisfies the following identity for any $(t, x) \in[0, T] \times \Omega$ :

$$
\begin{align*}
D_{t} v(t, x)= & \mathcal{A} v(t, x)+\left([h(\cdot)+E q(\cdot, \rho(x))] *\left[\mathcal{B} v(\cdot, x)+\mathcal{B} D_{t} u_{1}(\cdot, x)\right]\right)(t) \\
& +[h(t)+E q(t, \rho(x))] \mathcal{B} u_{0}(x)+\left(q(\cdot, \rho(x)) *\left[\mathcal{C} v(\cdot, x)+\mathcal{C} D_{t} u_{1}(\cdot, x)\right]\right)(t) \\
& +q(t, \rho(x)) \mathcal{C} u_{0}(x)+\left(\mathcal{A}-D_{t}\right) D_{t} u_{1}(t, x)+D_{t} f(t, x) \tag{19}
\end{align*}
$$

Moreover, evaluating (2) in $t=0$ and taking advantage of (4) we deduce that $v$ satisfies the initial condition

$$
\begin{equation*}
v(0, x)=\mathcal{A} u_{0}(x)+f(0, x)-D_{t} u_{1}(0, x)=: v_{0}(x), \quad \forall x \in \Omega, \tag{20}
\end{equation*}
$$

whereas, concerning the boundary values, from (14) and (5), (6) we get

$$
\begin{equation*}
v(t, \cdot)=D_{\nu} v(t, \cdot)=0, \quad \text { on } \partial \Omega, \quad \forall t \in[0, T] . \tag{21}
\end{equation*}
$$

Finally, since $\Phi$ and $\Psi$ act on spatial variables, only, they commute with the derivative with respect to time $D_{t}$. Consequently, from (10) and (11) we get

$$
\begin{align*}
& \Phi[v(t, \cdot)](\eta)=D_{t} g_{1}(t, \eta)-\Phi\left[D_{t} u_{1}(t, \cdot)\right](\eta), \quad \forall(t, \eta) \in[0, T] \times\left(l_{3}, l_{4}\right),  \tag{22}\\
& \Psi[v(t, \cdot)]=D_{t} g_{2}(t)-\Psi\left[D_{t} u_{1}(t, \cdot)\right], \quad \forall t \in[0, T] \tag{23}
\end{align*}
$$

Now, the consistency conditions related to problem (19)-(23) are explicitly given by:

$$
\begin{array}{llll}
(\mathrm{C} 2, \mathrm{D}, \mathrm{~N}) \quad v_{0}=0 & \text { on } \Gamma_{\mathrm{D}} \quad \text { and } & D_{\nu} v_{0}=0 & \text { on } \Gamma_{\mathrm{N}} \\
\Phi\left[v_{0}\right](\eta)=D_{t} g_{1}(0, \eta)-\Phi\left[D_{t} u_{1}(0, \cdot)\right](\eta), & \Psi\left[v_{0}\right]=D_{t} g_{2}(0)-\Psi\left[D_{t} u_{1}(0, \cdot)\right] \tag{25}
\end{array}
$$

Having derived the fundamental equations for $v$, we now turn our attention to $h$ and $q$. Applying $\Phi$ and $\Psi$ to both sides of (19) and taking advantage of (15) and (16), it is easy to check that, for any $(t, \eta) \in[0, T] \times\left(l_{3}, l_{4}\right)$, the following equations hold true:

$$
\begin{align*}
q(t, \eta) \Phi\left[\mathcal{C} u_{0}\right](\eta)+E q(t, \eta) \Phi\left[\mathcal{B} u_{0}\right](\eta)= & N_{1}^{0}\left(u_{1}, g_{1}, f\right)(t, \eta)+\Phi\left[N_{1}(v, h, q)(t, \cdot)\right](\eta) \\
& -\Phi_{1}[v(t, \cdot)](\eta)-h(t) \Phi\left[\mathcal{B} u_{0}\right](\eta)  \tag{26}\\
\Psi\left[q(t, \cdot) \mathcal{C} u_{0}+E q(t, \cdot) \mathcal{B} u_{0}\right]= & N_{2}^{0}\left(u_{1}, g_{2}, f\right)(t)+\Psi\left[N_{1}(v, h, q)(t, \cdot)\right] \\
& -\Psi_{1}[v(t, \cdot)]-h(t) \Psi\left[\mathcal{B} u_{0}\right] \tag{27}
\end{align*}
$$

where, for any $(t, \eta, x) \in[0, T] \times\left(l_{3}, l_{4}\right) \times \Omega$, operators $N_{1}^{0}, N_{2}^{0}$ and $N_{1}$ are defined, respectively, by formulae (3.2.19), (3.2.20) in [3] and

$$
\begin{align*}
N_{1}(v, h, q)(t, x)=- & \left([h(\cdot)+E q(\cdot, \rho(x))] *\left[\mathcal{B} v(\cdot, x)+\mathcal{B} D_{t} u_{1}(\cdot, x)\right]\right)(t) \\
& +\left(q(\cdot, \rho(x)) *\left[\mathcal{C} v(\cdot, x)+\mathcal{C} D_{t} u_{1}(\cdot, x)\right]\right)(t) \tag{28}
\end{align*}
$$

Remark 3.1. Now, suppose that the triplet $(v, h, q)$ solves the identification problem (19)-(23), (26) and (27). Then, performing integrations with respect to time and taking advantage of the consistency conditions (12), (13), it is easy to show that the pair $(u, k)$, related to $(v, h, q)$ by (9) and (14), solves our original problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$. Hence, problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$ and problem (19)-(23), (26), (27) are equivalent.

Now, due to (8), to determine $k(t, \cdot)$ for any $t \in[0, T]$ we have to solve system (26), (27) for $h$ and $q$. This part involves a lot of definition that will probably takes out of our purposes. Therefore, for brevity, here we limit ourselves to refer the reader to [3, Section 3.2] and to recall only that the pair $(h, q)$ solves the following fixed-point equations:

$$
\begin{align*}
& h(t)=h_{0}(t)+N_{3}(v, h, q)(t)  \tag{29}\\
& q(t, \eta)=q_{0}(t, \eta)+J_{2}\left(u_{0}\right)(\eta) N_{3}(v, h, q)(t)+N_{2}(v, h, q)(t, \eta) \tag{30}
\end{align*}
$$

Here $h_{0}$ and $q_{0}$ are defined, respectively, by formulae (3.2.38) and (3.2.42) in [3] whereas

$$
\begin{aligned}
& N_{2}(v, h, q)(t, \eta)=J_{3}\left(u_{0}\right)\left\{\Phi\left[N_{1}(v, h, q)(t, \cdot)\right](\eta)-\Phi_{1}[v(t, \cdot)](\eta)\right\}, \\
& N_{3}(v, h, q)(t):=\left[J_{1}\left(u_{0}\right)\right]^{-1}\left\{\Psi\left[N_{1}(v, h, q)(t, \cdot)\right]-\Psi\left[N_{2}(v, h, q)(t, \cdot) \mathcal{C} u_{0}\right]\right. \\
& \left.+\Psi\left[E\left(N_{2}(v, h, q)(t, \cdot)\right) \mathcal{B} u_{0}\right]-\Psi_{1}[v(t, \cdot)]\right\}, \\
& J_{2}\left(u_{0}\right)(\eta):=-\frac{\Phi\left[\mathcal{B} u_{0}\right](\eta)}{\Phi\left[\mathcal{C} u_{0}\right](\eta)} \exp \left[\int_{\eta}^{\eta_{0}} \frac{\Phi\left[\mathcal{B} u_{0}\right](\xi)}{\Phi\left[\mathcal{C} u_{0}\right](\xi)} \mathrm{d} \xi\right],
\end{aligned}
$$

the linear operator $J_{3}\left(u_{0}\right)$ appearing in $N_{2}$ being that defined by formula (3.2.30) in [3]. We can now summarize the result of this section in the following equivalence theorem.

Theorem 3.2. The pair $(u, k)$ is a solution to the identification problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$ if and only if the triplet $(v, h, q)$ defined by (9) and (14) solves the problem (19)-(23), (29), (30).

## 4. AN ABSTRACT REFORMULATION OF THE PROBLEM (19)-(23), (29), (30)

Starting from the result of Section 3, we now reformulate our identification problem in a Banach space framework. Let $X$ be a complex Banach space with norm $\|\cdot\|_{X}$ and let $A: \mathcal{D}(A) \subset X \rightarrow X$ be a linear operator, with a non-necessarily dense domain, satisfying the following assumption:
(H1) The resolvent $\rho(A)$ of $A$ contains the half-plane $S_{0}=\{\lambda \in \mathbf{C}: \operatorname{Re} \lambda \geq 0\}$ and there exists $M_{0}>0$ such that $\left\|(z I-A)^{-1}\right\|_{\mathcal{L}(X)} \leq M_{0}|1+z|^{-1}$ for every $z \in S_{0}$.
As it is well-known, assumption (H1) implies that $A$ generates an analytic semigroup $\left\{\mathrm{e}^{t A}\right\}_{t \geq 0}$ satisfying the estimates $\left\|A^{k} \mathrm{e}^{t A}\right\|_{\mathcal{L}(X)} \leq \widetilde{c}_{k} t^{-k}, t>0, k \in \mathbf{N} \cup\{0\}$. Moreover, from now on, we denote by $X_{1 / 2}$ the intermediate space $\mathcal{D}_{A}(1 / 2, p), p \in[1,+\infty]$ between $\mathcal{D}(A)$ and $X$ (cf. [8]).

In order to reformulate in an abstract form the problem (19)-(23), (29), (30) we need the following list of assumptions involving spaces, operators and data, where $0<\beta<\alpha<1 / 2$ and $\widetilde{q}_{0}$ is defined in Remark 3.3.3 in [3]:
(H2) $Y, Y_{1}, \mathcal{D}(B), \mathcal{D}(C)$ are Banach spaces such that $Y_{1} \hookrightarrow Y$ and $\mathcal{D}(A) \hookrightarrow \mathcal{D}(B) \hookrightarrow$ $\mathcal{D}(C) \hookrightarrow X, X_{1 / 2} \hookrightarrow \mathcal{D}(C) ;$
(H3) $B: \mathcal{D}(B) \rightarrow X$ and $C: \mathcal{D}(C) \rightarrow X$ are linear operators such that $B A^{-1} \in \mathcal{L}(X)$ and $C A^{-1} \in \mathcal{L}(X ; \mathcal{D}(C))$;
(H4) $E \in \mathcal{L}\left(Y ; Y_{1}\right), \Phi \in \mathcal{L}(X ; Y), \Phi_{1} \in \mathcal{L}(\mathcal{D}(C) ; Y), \Psi \in X^{*}, \Psi_{1} \in \mathcal{D}(C)^{*} ;$
(H5) $\mathcal{M}$ is a continuous bilinear operator from $Y \times \mathcal{D}(C)$ to $X$ and from $Y_{1} \times X$ to $X$;
$(\mathrm{H} 6) J_{1}: \mathcal{D}(B) \rightarrow \mathbf{R}, J_{2}: \mathcal{D}(B) \rightarrow Y, J_{3}: \mathcal{D}(B) \rightarrow \mathcal{L}(Y)$, are three prescribed (nonlinear) operators;

$$
\begin{aligned}
& \text { (H7) } u_{0} \in \mathcal{D}(B), C u_{0} \in \mathcal{D}(C), J_{1}\left(u_{0}\right) \neq 0, B u_{0} \in \mathcal{D}_{A}(\alpha,+\infty) \\
& \text { (H8) } v_{0} \in \mathcal{D}(A), h_{0} \in C^{\beta}([0, T] ; \mathbf{R}), q_{0} \in C^{\beta}([0, T] ; Y) ; \\
& \text { (H9) } z_{0} \in C^{\beta}([0, T] ; X), \quad z_{1} \in C^{\beta}([0, T] ; \mathcal{D}(C)), \quad z_{2} \in C^{\beta}([0, T] ; X) ; \\
& \text { (H10) } A v_{0}+\mathcal{M}\left(\widetilde{q}_{0}, C u_{0}\right)-\mathcal{M}\left(E \widetilde{q}_{0}, B u_{0}\right)+z_{2}(0, \cdot) \in \mathcal{D}_{A}(\beta,+\infty)
\end{aligned}
$$

Now (cf. (19)), denoting by $\mathcal{K}$ the convolution operator $\mathcal{K}(\chi, \kappa)(t):=\int_{0}^{t} \mathcal{M}(\chi(t-s), \kappa(s)) \mathrm{d} s$, our direct problem depending on the pair of parameter $(h, q)$, is the following: determine a function $v \in$ $C^{1}([0, T] ; X) \cap C([0, T] ; \mathcal{D}(A))$ satisfying

$$
\left\{\begin{align*}
v^{\prime}(t)= & {\left[\lambda_{0} I+A\right] v(t)+\left[h *\left(B v+z_{0}\right)\right](t)-\mathcal{K}\left(E q, B v+z_{0}\right)(t)+\mathcal{M}\left(q(t), C u_{0}\right) }  \tag{31}\\
& +h(t) B u_{0}+\mathcal{K}\left(q, C v+z_{1}\right)(t)-\mathcal{M}\left(E q(t), B u_{0}\right)+z_{2}(t), \quad \forall t \in[0, T] \\
v(0)= & v_{0}
\end{align*}\right.
$$

Remark 4.1. In the explicit case (19), we have $A=\mathcal{A}-\lambda_{0} I$, with a large enough positive $\lambda_{0}$, and $z_{0}=\mathcal{B} D_{t} u_{1}, z_{1}=\mathcal{C} D_{t} u_{1}, z_{2}=\left(\mathcal{A}-D_{t}\right) D_{t} u_{1}+D_{t} f$. Instead, functions $v_{0}, h_{0}$ and $q_{0}$ in (H8) are those appearing, respectively, in (20), (29) and (30).

We now introduce the following unknown function $w$ :

$$
w=A v \quad \Longleftrightarrow \quad v=A^{-1} w
$$

Hence, applying $A$ to the Volterra operator equation equivalent to problem (31), we obtain that $w$ solves the equation

$$
\begin{equation*}
w=w_{0}+R_{1}(w, h, q)+S_{1}(q) \tag{32}
\end{equation*}
$$

where we have set

$$
\begin{aligned}
& \begin{aligned}
w_{0}:=A \mathrm{e}^{t A} v_{0} & +A\left(\mathrm{e}^{t A} * z_{2}\right) \\
R_{1}(w, h, q):= & \lambda_{0}\left(\mathrm{e}^{t A} * w\right)+A\left[\mathrm{e}^{t A} * h *\left(B A^{-1} w+z_{0}\right)\right]+A\left(h * \mathrm{e}^{t A} B u_{0}\right) \\
& \quad-A\left[\mathrm{e}^{t A} * \mathcal{K}\left(E q, B A^{-1} w+z_{0}\right)\right]+A\left[\mathrm{e}^{t A} * \mathcal{K}\left(q, C A^{-1} w+z_{1}\right)\right]
\end{aligned} \\
& S_{1}(q):=A\left[\mathrm{e}^{t A} *\left(\mathcal{M}\left(q, C u_{0}\right)-\mathcal{M}\left(E q, B u_{0}\right)\right)\right] .
\end{aligned}
$$

Now we rewrite the fixed-point system (29), (30) in an abstract form. For this purpose we introduce the operators:

$$
\begin{aligned}
& R_{2}(w, h, q):=-\widetilde{J_{1}}\left(u_{0}\right)\{\Psi[\mathcal{M}( J_{3}\left(u_{0}\right) \Phi[ \\
&\left.\left.\quad-\mathcal{M}\left(E A^{-1} w, h, q\right)\right], C u_{0}\right)-N_{1}\left(u_{0}\right) \Phi\left[N_{1}\left(A^{-1} w, h, q\right)\right. \\
&\left.\left.\left.\quad-B u_{0}\right)\right]\right\} \\
& R_{3}(w, h, q):=J_{2}\left(u_{0}\right) R_{2}(w, h, q)+J_{3}\left(u_{0}\right) \Phi\left[N_{1}\left(A^{-1} w, h, q\right)\right] \\
& S_{2}(w):=\widetilde{J_{1}}\left(u_{0}\right)\left\{\Psi\left[\mathcal{M}\left(J_{3}\left(u_{0}\right) \Phi_{1}\left[A^{-1} w\right], C u_{0}\right)-\mathcal{M}\left(E J_{3}\left(u_{0}\right) \Phi_{1}\left[A^{-1} w\right], B u_{0}\right)\right]-\Psi_{1}\left[A^{-1} w\right]\right\}, \\
& S_{3}(w):=J_{2}\left(u_{0}\right) S_{2}(w)-J_{3}\left(u_{0}\right) \Phi_{1}\left[A^{-1} w\right]
\end{aligned}
$$

where (cf. (H7) and (28)) we have set $\widetilde{J_{1}}\left(u_{0}\right)=\left[J_{1}\left(u_{0}\right)\right]^{-1}$ and

$$
N_{1}\left(A^{-1} w, h, q\right)=-h *\left(B A^{-1} w+z_{0}\right)+\mathcal{K}\left(E q, B A^{-1} w+z_{0}\right)-\mathcal{K}\left(q, C A^{-1} w+z_{1}\right)
$$

Then, the fixed-point system for $h$ and $q$ can be rewritten in the following more compact form:

$$
\begin{equation*}
h=h_{0}+R_{2}(w, h, q)+S_{2}(w), \quad q=q_{0}+R_{3}(w, h, q)+S_{3}(w) \tag{33}
\end{equation*}
$$

$h_{0}$ and $q_{0}$ being the elements appearing in (H8). Now, the fixed-point system (32) and (33) coincide with that in Section 5 of [2]. Therefore, referring the reader to Section 6 of [2] for the proof, we can state the following local in time existence and uniqueness result.

Theorem 4.2. Under assumptions (H1)-(H10) there exists $T^{*} \in(0, T)$ such that for any $\tau \in\left(0, T^{*}\right]$ the fixed-point system (32), (33) has a unique solution $(w, h, q) \in C^{\beta}([0, \tau] ; X) \times C^{\beta}([0, \tau] ; \mathbf{R}) \times C^{\beta}([0, \tau] ; Y)$.

Corollary 4.3. Under assumptions (H1)-(H10) there exists $T^{*} \in(0, T)$ such that for any $\tau \in\left(0, T^{*}\right]$ the problem (31), (33) admits a unique solution $(v, h, q) \in\left[C^{1+\beta}([0, \tau] ; X) \cap C^{\beta}([0, \tau] ; \mathcal{D}(A))\right] \times C^{\beta}([0, \tau] ; \mathbf{R}) \times$ $C^{\beta}([0, \tau] ; Y)$.

## 5. APPLICATION OF THE ABSTRACT RESULT TO P(D,N)

In this section, taking advantage of the equivalence Theorem 3.2 and the abstract Corollary 4.3, we solve (locally in time) the identification problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$, at least with the convention to work in the framework of Sobolev spaces related to $L^{p}(\Omega)$ with

$$
\begin{equation*}
p \in(3,+\infty) \tag{34}
\end{equation*}
$$

We start by listing our main requirements on the operators $\Phi$ and $\Psi$. They are the same as in Section 3, but here we rewrite them in a more formal way, making clear the minimal space regularity required for the functions $v_{j}, j=1,2,3$, and the operators $\Phi_{1}, \Psi_{1}$ appearing on (15) and (16), respectively. Recalling that $\Xi$ is the set of spatial variables on which $\Phi$ acts and taking into account (7), we assume:

$$
\begin{align*}
& \Phi \in \mathcal{L}\left(L^{p}(\Omega) ; L^{p}\left(l_{3}, l_{4}\right)\right), \quad \Psi \in L^{p}(\Omega)^{*},  \tag{35}\\
& \Phi\left[v_{1} v_{2}\right]=v_{1} \Phi\left[v_{2}\right], \quad \forall\left(v_{1}, v_{2}\right) \in L^{p}\left(l_{3}, l_{4}\right) \times L^{p}(\Omega),  \tag{36}\\
& D_{\eta} \Phi\left[v_{3}\right](\eta)=\Phi\left[D_{\eta} v_{3}\right](\eta), \quad \forall v_{3} \in W^{1, p}\left(\left(l_{3}, l_{4}\right) \times \Xi\right) \text { and } \eta \in\left(l_{3}, l_{4}\right),  \tag{37}\\
& \Phi \mathcal{A}=\mathcal{A}_{1} \Phi+\Phi_{1} \text { on } W_{\mathrm{D}, \mathrm{~N}}^{2, p}(\Omega), \quad \Phi_{1} \in \mathcal{L}\left(W^{1, p}(\Omega) ; L^{p}\left(l_{3}, l_{4}\right)\right),  \tag{38}\\
& \Psi \mathcal{A}=\Psi_{1} \text { on } W_{\mathrm{D}, \mathrm{~N}}^{2, p}(\Omega), \quad \Psi_{1} \in W^{1, p}(\Omega)^{*} \tag{39}
\end{align*}
$$

where $\mathcal{A}_{1}$ is a second-order differential operator $\mathcal{A}_{1}=\mathcal{A}_{1}\left(D_{\eta}\right)$ and $W_{\mathrm{D}, \mathrm{N}}^{k, s}(\Omega), k \in \mathbf{N}, s \in[1,+\infty]$, denote the space of function $\omega \in W^{k, s}(\Omega)$ satisfying the homogeneous condition (D,N).

To state our result concerning the identification problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$ we need to list also the following assumptions on the data $f, u_{j}, g_{k}, j=0,1, k=1,2$ :

$$
\begin{align*}
& f \in C^{1+\beta}\left([0, T] ; L^{p}(\Omega)\right), \quad f(0, \cdot) \in W^{2, p}(\Omega)  \tag{40}\\
& u_{1} \in C^{2+\beta}\left([0, T] ; L^{p}(\Omega)\right) \cap C^{1+\beta}\left([0, T] ; W^{2, p}(\Omega)\right),  \tag{41}\\
& u_{0} \in W^{4, p}(\Omega), \quad v_{0} \in W_{\mathrm{D}, \mathrm{~N}}^{2, p}(\Omega), \quad \mathcal{B} u_{0} \in W_{\mathrm{D}, \mathrm{~N}}^{2 \alpha, p}(\Omega),  \tag{42}\\
& F:=k_{0}^{\prime} \mathcal{C} u_{0}+k_{0} \mathcal{B} u_{0}+\mathcal{A} v_{0}+\left(\mathcal{A}-D_{t}\right) D_{t} u_{1}(0, \cdot)+D_{t} f(0, \cdot) \in W_{\mathrm{D}, \mathrm{~N}}^{2 \beta, p}(\Omega),  \tag{43}\\
& g_{1} \in C^{2+\beta}\left([0, T] ; L^{p}\left(l_{3}, l_{4}\right)\right) \cap C^{1+\beta}\left([0, T] ; W^{2, p}\left(l_{3}, l_{4}\right)\right),  \tag{44}\\
& \mathcal{A}_{1} D_{t} g_{1} \in C^{\beta}\left([0, T] ; L^{p}\left(l_{3}, l_{4}\right)\right),  \tag{45}\\
& g_{2} \in C^{2+\beta}([0, T] ; \mathbf{R}), \tag{46}
\end{align*}
$$

where $0<\beta<\alpha<1 / 2, \alpha, \beta \neq 1 /(2 p)$, and functions $v_{0}$ and $k_{0}$ appearing in (42) and (43) are defined, respectively, by formulae (20) and (3.18) in [4] (with the triplet $\left(r, R_{2}, \eta\right)$ being replaced by $\left(\eta, \eta_{0}, \sigma\right)$ ). Here the spaces $W_{\mathrm{D}, \mathrm{N}}^{2 \gamma, p}(\Omega) \equiv\left(L^{p}(\Omega), W_{\mathrm{D}, \mathrm{N}}^{2, p}(\Omega)\right)_{\gamma, p}, \gamma \in(0,1 / 2] \backslash\{1 /(2 p)\}$, are interpolation spaces between $W_{\mathrm{D}, \mathrm{N}}^{2, p}(\Omega)$ and $L^{p}(\Omega)$ and they are defined (cf. [11, Section 4.3.3]) by:

$$
W_{\mathrm{D}, \mathrm{~N}}^{2 \gamma, p}(\Omega)=\left\{\begin{array}{l}
W^{2 \gamma, p}(\Omega), \quad \text { if } \quad 0<\gamma<1 /(2 p),  \tag{47}\\
\left\{u \in W^{2 \gamma, p}(\Omega): \quad u=0 \quad \text { on } \Gamma_{\mathrm{D}}\right\}, \quad \text { if } \quad 1 /(2 p)<\gamma \leq 1 / 2, \\
W^{2 \gamma, p}(\Omega), \quad \text { if } \quad 0<\gamma \leq 1 / 2 \text { and } \quad \Gamma_{\mathrm{D}}=\emptyset
\end{array}\right.
$$

Further, we introduce the Banach spaces $\mathcal{U}^{s, \beta, p}(T), s \in \mathbf{N} \backslash\{0\}$, defined by

$$
\begin{equation*}
\mathcal{U}^{s, \beta, p}(T)=C^{s+\beta}\left([0, T] ; L^{p}(\Omega)\right) \cap C^{s-1+\beta}\left([0, T] ; W^{2, p}(\Omega)\right) . \tag{48}
\end{equation*}
$$

Analogously, replacing $W^{2, p}(\Omega)$ with $W_{\mathrm{D}, \mathrm{N}}^{2, p}(\Omega)$ in (48), we define spaces $\mathcal{U}_{\mathrm{D}, \mathrm{N}}^{s, \beta, p}(T)$.
Finally, we assume that the coefficients of operators $\mathcal{A}, \mathcal{B}, \mathcal{C}$ satisfy the following properties:

$$
\begin{equation*}
a_{i, j} \in W^{2, \infty}(\Omega), \quad a_{i, j}=a_{j, i}, \quad b_{i, j}, c_{i} \in W^{1, \infty}(\Omega), \quad i, j=1,2,3 \tag{49}
\end{equation*}
$$

We can now state the main result of the section.
Theorem 5.1. Let assumptions (3), (49), (34)-(39) be fulfilled and assume the function $\rho$ satisfies the assumptions (A)-(E) of Section 2. Moreover, let the data vector $\left(u_{0}, u_{1}, g_{1}, g_{2}, f\right)$ satisfies assumptions (40)-(46), inequalities (17), (18) and consistency conditions (12), (13), as well as (24), (25).

Then, there exists $T^{*} \in(0, T]$ such that the identification problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$ admits a unique solution $(u, k) \in \mathcal{U}^{2, \beta, p}\left(T^{*}\right) \times C^{\beta}\left(\left[0, T^{*}\right] ; W^{1, p}\left(l_{3}, l_{4}\right)\right)$ depending continuously on the data with respect to the norms related to the Banach spaces in (40)-(46).

To prove Theorem 5.1 we take advantage of the equivalence result of Section 3. Indeed, in the space framework above defined, Theorem 3.2 reads as follows.

Theorem 5.2. The pair $(u, k) \in \mathcal{U}^{2, \beta, p}\left(T^{*}\right) \times C^{\beta}\left(\left[0, T^{*}\right] ; W^{1, p}\left(l_{3}, l_{4}\right)\right)$ is a solution to the identification problem $\mathrm{P}(\mathrm{D}, \mathrm{N})$ if and only if the triplet $(v, h, q)$ defined by (9) and (14) belongs to $\mathcal{U}_{\mathrm{D}, \mathrm{N}}^{1, \beta, p}\left(T^{*}\right) \times$ $C^{\beta}\left(\left[0, T^{*}\right] ; \mathbf{R}\right) \times C^{\beta}\left(\left[0, T^{*}\right] ; L^{p}\left(l_{3}, l_{4}\right)\right)$ and solves the problem (19)-(23), (29), (30).

Consequently, it suffices to prove the following theorem, because then Theorem 5.1 trivially follows from Theorem 5.2.

Theorem 5.3. Let the assumptions of Theorem 5.2 be fulfilled. Then, the identification problem (19)(23), (29), (30) admits a unique solution $(v, h, q) \in \mathcal{U}_{\mathrm{D}, \mathrm{N}}^{1, p}\left(T^{*}\right) \times C^{\beta}\left(\left[0, T^{*}\right] ; \mathbf{R}\right) \times C^{\beta}\left(\left[0, T^{*}\right] ; L^{p}\left(l_{3}, l_{4}\right)\right)$ depending continuously on the data with respect to the norms related to the Banach spaces in (40)-(46).

In order to prove Theorem 5.3 a preliminary result is needed. It deals with an estimate which, in the proof of Theorem 5.3, will allow us to prove assumption (H5) for the bilinear operator:

$$
\mathcal{M}:\left\{\begin{array}{l}
L^{p}\left(l_{3}, l_{4}\right) \times W^{1, p}(\Omega) \\
W^{1, p}\left(l_{3}, l_{4}\right) \times L^{p}(\Omega)
\end{array} \longrightarrow L^{p}(\Omega), \quad \mathcal{M}(q, w)(x)=q(\rho(x)) w(x), \quad x \in \Omega\right.
$$

Lemma 5.4. Let $\Omega \subset \mathbf{R}^{3}$ be a bounded domain satisfying the assumptions of Section 2 and let $\rho$ be a function satisfying the assumptions (A)-(E) of the same section. Then, for every $s \in[1,+\infty)$, the following estimate holds

$$
\begin{equation*}
\|g \circ \rho\|_{L^{s}(\Omega)} \leq C_{1}(\rho, \Omega)\|g\|_{L^{s}\left(l_{3}, l_{4}\right)}, \quad \forall g \in L^{s}\left(l_{3}, l_{4}\right) \tag{50}
\end{equation*}
$$

where $C_{1}(\rho, \Omega)$ is a positive constant depending on $\rho$ and $\Omega$, only.
Proof. By recalling $\bar{\Omega} \subset \bar{V}$, we set $\Omega_{1}=\bar{\Omega} \backslash \partial V$ and $\Omega_{2}=\bar{\Omega} \cap \partial V$, so that $\Omega_{1} \subset V, \Omega_{2} \subset \partial V$ and $\bar{\Omega}=\Omega_{1} \cup \Omega_{2}$. Now, since $|\nabla \rho(\cdot)|$ is positive in $V$, for every $y \in \Omega_{1}$ there exists $k=k(y) \in\{1,2,3\}$ such that $D_{x_{k}} \rho(y) \neq 0$. According to this, for every $y \in \Omega_{1}$, we denote by $i(y)$ the minimum index $k \in\{1,2,3\}$ such that $D_{y_{k}} \rho(y) \neq 0$ and by $B_{y}^{i(y)}$ the ball $B\left(y, r_{y}\right), r_{y}>0$, such that $D_{x_{i(y)}} \rho(z) \neq 0$ for every $z \in \overline{B\left(y, r_{y}\right)}$. This is possible since $\Omega_{1} \subset V$ and $V$ is open. Hence, at least by taking sufficiently small $r_{y}$, the ball $B_{y}^{i(y)}$ is well defined. Similarly (cf. (D)), for every $y \in \Omega_{2}$ there exist $r_{y}>0$, $i(y)=\min \{1,2,3\}$ and a subset $U_{y}^{i(y)}$ of $\mathbf{R}^{3}$ having zero three-dimensional Lebesgue measure such that $D_{x_{i(y)}} \rho(z) \neq 0$ for every $z \in B\left(y, r_{y}\right) \backslash U_{y}^{i(y)}$. Denoting with $\widetilde{B}_{y}^{i(y)}$ the ball $B\left(y, r_{y}\right), y \in \partial V$, such that (D) is satisfied, it follows that the collection $\left\{B_{y}^{i(y)}\right\}_{y \in \Omega_{1}}$ and $\left\{\widetilde{B}_{y}^{i(y)}\right\}_{y \in \Omega_{2}}$ cover $\Omega_{1} \backslash \bigcup_{y \in \Omega_{2}} \widetilde{B}_{y}^{i(y)}$ and $\Omega_{2}$, respectively. As a consequence

$$
\bar{\Omega}=\Omega_{1} \cup \Omega_{2} \subset\left(\bigcup_{y \in \Omega_{1}} B_{y}^{i(y)}\right) \cup\left(\bigcup_{y \in \Omega_{2}} \widetilde{B}_{y}^{i(y)}\right)
$$

and hence, via a compactness argument, we find $y_{1}, \ldots, y_{n_{1}} \in \Omega_{1}, y_{n_{1}+1}, \ldots, y_{n_{1}+n_{2}} \in \Omega_{2}, n_{j} \in \mathbf{N}$, $j=1,2$, such that:

$$
\begin{equation*}
\bar{\Omega} \subset\left[\left(\bigcup_{j=1}^{n_{1}} B_{y_{j}}^{i\left(y_{j}\right)}\right) \cup\left(\bigcup_{k=1}^{n_{2}} \widetilde{B}_{y_{n_{1}+k}}^{i\left(y_{n_{1}+k}\right)}\right)\right] \tag{51}
\end{equation*}
$$

Now, for $\nu=1,2,3, j=1, \ldots, n_{1}, k=1, \ldots, n_{2}$, we set $B^{\nu}=\bigcup_{i\left(y_{j}\right)=\nu} B_{y_{j}}^{i\left(y_{j}\right)}, \widetilde{B}^{\nu}=\bigcup_{i\left(y_{n_{1}+k}\right)=\nu} \widetilde{B}_{y_{n_{1}+k} i\left(y_{n_{1}+k}\right)}$, so that (51) can be rewritten as

$$
\begin{equation*}
\bar{\Omega} \subset\left[\left(\bigcup_{\nu=1}^{3} B^{\nu}\right) \cup\left(\bigcup_{\nu=1}^{3} \widetilde{B}^{\nu}\right)\right] \tag{52}
\end{equation*}
$$

From the definition of $B_{y}^{k}, y \in \Omega_{1}$, and $\widetilde{B}_{y}^{k}, y \in \Omega_{2}, k \in\{1,2,3\}$, for any fixed $\nu \in\{1,2,3\}$ we have $D_{x_{\nu}} \rho(z) \neq 0$ for every $z \in \bar{B}^{\nu} \cup\left(\widetilde{B}^{\nu} \backslash \bigcap_{i\left(y_{n_{1}+k}\right)=\nu, k=1, \ldots, n_{2}} U_{y_{n_{1}+k}}^{i\left(y_{n_{1}+k}\right)}\right)$, where $U_{y_{n_{1}+k}}^{i\left(y_{n_{1}+k}\right)}$ is the subset of three-dimensional Lebesgue measure equal to zero appearing on (D) and related to $y_{n_{1}+k}, k=1, \ldots, l_{2}$. In particular, the last assertion follows from the inclusion

$$
\bigcup_{j=1}^{n_{0}}\left(A_{j} \backslash B_{j}\right) \subset\left(\bigcup_{j=1}^{n_{0}} A_{j}\right) \backslash\left(\bigcap_{j=1}^{n_{0}} B_{j}\right)
$$

which is satisfied for any collection of sets $\left\{A_{j}\right\}_{j=1}^{n_{0}}$ and $\left\{B_{j}\right\}_{j=1}^{n_{0}}, n_{0} \in \mathbf{N}$.
For the sake of brevity, from now on, for every $\nu=1,2,3$, we denote by $U^{\nu}$ the subset (of threedimensional Lebesgue measure equal to zero) $\bigcap_{i\left(y_{n_{1}+k}\right)=\nu, k=1, \ldots, n_{2}} U_{y_{n_{1}+k}}^{i\left(y_{n_{1}+k}\right)}$. Therefore, for every $g \in$ $L^{s}\left(l_{3}, l_{4}\right)$, from (52) we get

$$
\begin{align*}
\|g \circ \rho\|_{L^{s}(\Omega)}^{s} \leq & \sum_{\nu=1}^{3}\left(\|g \circ \rho\|_{L^{s}\left(\Omega \cap B^{\nu}\right)}^{s}+\|g \circ \rho\|_{L^{s}\left(\Omega \cap \widetilde{B}^{\nu}\right)}^{s}\right) \\
& =\sum_{\nu=1}^{3}\left(\|g \circ \rho\|_{L^{s}\left(\Omega \cap B^{\nu}\right)}^{s}+\|g \circ \rho\|_{L^{s}\left(\Omega \cap\left(\widetilde{B}^{\nu} \backslash U^{\nu}\right)\right)}^{s}\right) . \tag{53}
\end{align*}
$$

Let now $\nu \in\{1,2,3\}$ be fixed and, for $j, k \in\{1,2,3\} \backslash\{\nu\}, j \neq k$, consider the change of variables $F^{\nu}$ : $\left(x_{\nu}, x_{j}, x_{k}\right) \rightarrow\left(\xi_{\nu}, \xi_{j}, \xi_{k}\right)$, where $\left(x_{\nu}, x_{j}, x_{k}\right) \in \Omega \cap\left(B^{\nu} \cup\left(\widetilde{B}^{\nu} \backslash U^{\nu}\right)\right)$ and $\left(\xi_{\nu}, \xi_{j}, \xi_{k}\right)=\left(\rho\left(x_{\nu}, x_{j}, x_{k}\right), x_{j}, x_{k}\right)$. Since $D_{x_{\nu}} \rho(z) \neq 0$ for every $z \in B^{\nu} \cup\left(\widetilde{B}^{\nu} \backslash U^{\nu}\right)$, from the inverse function theorem we get $\left(x_{\nu}, x_{j}, x_{k}\right)=$ $\left(\lambda_{\nu}\left(\xi_{\nu}, \xi_{j}, \xi_{k}\right), \xi_{j}, \xi_{k}\right)=: G^{\nu}\left(\xi_{1}, \xi_{2}, \xi_{3}\right)$ for every $\left(\xi_{\nu}, \xi_{j}, \xi_{k}\right) \in F^{\nu}\left(\Omega \cap\left(B^{\nu} \cup\left(\widetilde{B}^{\nu} \backslash U^{\nu}\right)\right)\right)$. Of course, the set $F^{\nu}\left(\Omega \cap\left(B^{\nu} \cup\left(\widetilde{B}^{\nu} \backslash U^{\nu}\right)\right)\right.$ ) is not easily characterized, but we can surely say that it is contained in $\left[l_{3}, l_{4}\right] \times \Pi_{j, k}(\bar{\Omega})$, where $\Pi_{j, k}(\bar{\Omega}), j, k=1,2,3, j \neq k$, denotes the continuous projection of $\bar{\Omega}$ on the plane $x_{j} x_{k}$ acting in the following way:

$$
\Pi_{j, k}\left(x_{\nu}, x_{j}, x_{k}\right)=\left(0, x_{j}, x_{k}\right)
$$

In particular, $\bar{\Omega}$ being compact, $\Pi_{j, k}(\bar{\Omega}), j, k=1,2,3$, is a compact (and hence measurable) subset of $\mathbf{R}^{2}$ having two-dimensional Lebesgue measure $m_{2}\left(\Pi_{j, k}(\bar{\Omega})\right)$ bounded from above by $m_{2}\left(\Pi_{j, k}(Q)\right), Q$ being any three-dimensional cube containing $\bar{\Omega}$. Therefore, denoting by $A(l)$ the set $\Pi_{j, k}(\{x \in \bar{\Omega} \cap V: \rho(x)=l\})$, $l \in\left[l_{3}, l_{4}\right], A(l)$ turns out to be a compact subset of $\Pi_{j, k}(\bar{\Omega})$, since the level set $\{x \in \bar{\Omega} \cap V: \rho(x)=l\}$, $l \in\left[l_{3}, l_{4}\right]$, is compact and $\Pi_{j, k}$ is a continuous map. As a consequence $A(l), l \in\left[l_{3}, l_{4}\right]$, is a twodimensional Lebesgue measurable subset of $\Pi_{j, k}(\bar{\Omega})$ and hence, using (E) and denoting by $\mathcal{J} F^{\nu}$ the Jacobian of $F^{\nu}$, from Fubini's theorem we get

$$
\begin{align*}
& \|g \circ \rho\|_{L^{s}\left(\Omega \cap B^{\nu}\right)}^{s}+\|g \circ \rho\|_{L^{s}\left(\Omega \cap\left(\widetilde{B}^{\nu} \backslash U^{\nu}\right)\right)}^{s} \\
= & \int_{l_{3}}^{l_{4}}\left|g\left(\xi_{\nu}\right)\right|^{s} \mathrm{~d} \xi_{\nu} \int_{A\left(\xi_{\nu}\right)}\left|\operatorname{det} \mathcal{J} F^{\nu}\left(G^{\nu}\left(\xi_{\nu}, \xi_{j}, \xi_{k}\right)\right)\right| \mathrm{d} \xi_{j} \mathrm{~d} \xi_{k} \leq \widetilde{C} m_{2}\left(A\left(\xi_{\nu}\right)\right)\|g\|_{L^{s}\left(l_{3}, l_{4}\right)}^{s} \\
\leq & \widetilde{C} m_{2}\left(\Pi_{j, k}(\bar{\Omega})\right)\|g\|_{L^{s}\left(l_{3}, l_{4}\right)}^{s}, \quad \forall g \in L^{s}\left(l_{3}, l_{4}\right) . \tag{54}
\end{align*}
$$

Now, since $\nu$ was any index belonging to $\{1,2,3\}$, by replacing (54) in (53) and summing up on $\nu$ we obtain (50). The proof is complete.

With the Lemma 5.4 at hand, we can now prove Theorem 5.3.
Proof of Theorem 5.3. First, for any $p \in(3,+\infty)$ let us choose the Banach spaces $X, X_{1 / 2}, \mathcal{D}(A), \mathcal{D}(B)$, $\mathcal{D}(C), Y$ and $Y_{1}$ according to the rule

$$
\left\{\begin{array}{l}
X=L^{p}(\Omega), \quad X_{1 / 2}=W_{\mathrm{D}, \mathrm{~N}}^{1, p}(\Omega)  \tag{55}\\
\mathcal{D}(A)=W_{\mathrm{D}, \mathrm{~N}}^{2, p}(\Omega), \quad \mathcal{D}(B)=W^{2, p}(\Omega), \quad \mathcal{D}(C)=W^{1, p}(\Omega) \\
Y=L^{p}\left(l_{3}, l_{4}\right), \quad Y_{1}=W^{1, p}\left(l_{3}, l_{4}\right)
\end{array}\right.
$$

where the spaces $W_{\mathrm{D}, \mathrm{N}}^{1, p}(\Omega)$ are defined by (47) with $\gamma=1 / 2$. Then, $\mathcal{A}, \mathcal{B}, \mathcal{C}$ being defined by (3) and $\lambda_{0}$ being a large enough (fixed) positive constant, we define the operators $A, B, C$ with domains $\mathcal{D}(A)$, $\mathcal{D}(B), \mathcal{D}(C)$ as follows

$$
\begin{equation*}
A u=\left(\mathcal{A}-\lambda_{0} I\right) u, u \in \mathcal{D}(A) ; \quad B u=\mathcal{B} u, u \in \mathcal{D}(B) ; \quad C u=\mathcal{C} u, u \in \mathcal{D}(C) \tag{56}
\end{equation*}
$$

(55) and (56) ensure that (H2) is satisfied, since (cf. [8]) $X_{1 / 2}=(X, \mathcal{D}(A))_{1 / 2, p}=\mathcal{D}_{A}(1 / 2, p)$. For the brevity's sake, we refer to [3, pagg. 87-90] for the proof that (H1), (H3), (H4) and (H6)-(H10) are satisfied, and, instead, we turn our attention to (H5) where Lemma 5.4 plays a fundamental role. Indeed, to show that $\mathcal{M}$ is a continuous bilinear operator from $L^{p}\left(l_{3}, l_{4}\right) \times W^{1, p}(\Omega)$ to $L^{p}(\Omega)$ and from $W^{1, p}\left(l_{3}, l_{4}\right) \times L^{p}(\Omega)$ to $L^{p}(\Omega)$ we proceed as follows. First, using Lemma 5.4 and the Sobolev embedding $W^{1, p}(\Omega) \hookrightarrow C^{1-3 / p}(\bar{\Omega}), p \in(3,+\infty)$, for every pair $(q, w) \in L^{p}\left(l_{3}, l_{4}\right) \times W^{1, p}(\Omega)$ we find:

$$
\|\mathcal{M}(q, w)\|_{L^{p}(\Omega)} \leq\|q \circ \rho\|_{L^{p}(\Omega)}\|w\|_{C(\bar{\Omega})} \leq C_{3}(\rho, \Omega, p)\|q\|_{L^{p}\left(l_{3}, l_{4}\right)}\|w\|_{W^{1, p}(\Omega)}
$$

where the positive constant $C_{3}(\rho, \Omega, p)$ depends on $\rho, \Omega$ and $p$, only, and is given by the product of the constant $C_{1}(\rho, \Omega)$ in (50) with the constant $C_{2}(p, \Omega)$ such that $\|w\|_{C(\bar{\Omega})} \leq C_{2}(p, \Omega)\|w\|_{W^{1, p}(\Omega)}$. Similarly, using $W^{1, p}\left(l_{3}, l_{4}\right) \hookrightarrow C^{1-1 / p}\left(\left[l_{3}, l_{4}\right]\right), p \geq 1$, for any pair $(q, w) \in W^{1, p}\left(l_{3}, l_{4}\right) \times L^{p}(\Omega)$ we get:

$$
\|\mathcal{M}(q, w)\|_{L^{p}(\Omega)} \leq C_{4}(p, \Omega)\|q\|_{W^{1, p}\left(l_{3}, l_{4}\right)}\|w\|_{L^{p}(\Omega)} .
$$

Hence, the bilinearity of $\mathcal{M}$ being obvious, we have that assumption (H5) is satisfied.

## References

1. D. S. Cohen and A. B. White Jr., Sharp fronts due to diffusion and viscoelastic relaxation in polymers. SIAM J. Appl. Math. (1991) 51, 472-483.
2. F. Colombo and A. Lorenzi, An identification problem related to parabolic integrodifferential equations with non commuting spatial operators. J. Inverse Ill-Posed Probl. (2000) 8, 505-540.
3. A. Favaron, Recovering Conductivity Coefficients and Memory Kernels in Parabolic Integro-Differential Equations: Applications to Domain with Special Symmetries, PhD. Thesis, Dipartimento di Matematica, Bologna, 2004.
4. A. Favaron and A. Lorenzi, Parabolic integro-differential identification problems related to radial memory kernels I. J. Inverse Ill-Posed Probl. (2001) 9, 489-529.
5. J. Janno, An inverse problem arising in compression of visco-elastic medium. Proc. Estonian Acad. Sci. Phys. Math. (2000) 49, 75-89.
6. J. Janno, Inverse problems for determining monotone weakly singular relaxation kernels in viscoelasticity. Nonlinear Analysis (2000) 41, 943-962.
7. J. Janno and L. von Wolfersdorf, An inverse problem for identification of a time- and space-dependent memory kernel of a special kind in heat conduction. Inv. Prob. (1999) 15, 1455-1467.
8. A. Lunardi, Analytic Semigroups and Optimal Regularity in Parabolic Problems, Birkhäuser Verlag, Basel, 1995.
9. M. Renardy, W. J. Hrusa and J. A. Nohel, Mathematical Problems in Viscoelasticity, Longman Scientific \& Technical, Harlow, 1987.
10. S. Shaw and J. R. Whiteman, Some partial differential Volterra equation problems arising in viscoelasticity, BICOM Tech. Rep. 97/9, http://people.brunel.ac.uk/~icsrsss/bicom/tech_rep.html\#97.
11. H. Triebel, Interpolation Theory, Function Spaces, Differential Operators, North Holland Publ. Co., Amsterdam, 1978.
