# RECONSTRUCTING TRAJECTORIES FROM MEASUREMENT DATA $\dagger$ 

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Practical problems sonaetimes require the reconstruction of the continuous flight trajectory of an object from the results of coordinate measurements performed continuously or at discrete times, with errors. Here it is required to minimize the deviation of the reconstructed trajectory from the measured trajectory, also using available restrictions on the linear acceleration of the object. This paper presents a statement of the problem and a method of solving it. An example is given.

## 1. STATEMENT OF THE PROBLEM

Suppose that a material point is acted upon by a controlling acceleration $u(t)$ so that it performs onedimensional motion described by the equations

$$
\begin{equation*}
\dot{x}=y, \quad \dot{y}=u \tag{1.1}
\end{equation*}
$$

where $x$ is the coordinate and $y$ the velocity of the point. The acceleration $u(t)$ is constrained by the condition

$$
\begin{equation*}
|u(t)| \leqslant D \tag{1.2}
\end{equation*}
$$

where $D$ is a constant. Measurements of the coordinate $x(t)$ are performed during the interval $t \in[0, T]$, the results of the measurements being denoted by $h(t)$. It is required to reconstruct the trajectory of the object, i.e. to find a function $x(t)$ in the interval $t \in[0, T]$ so that constraint (1.2) is satisfied and the deviation $x(t)-h(t)$ is in some sense minimized.
The problem can be formalized as an optimal control problem in the following way.
It is required to determine the control $u(t)$ which minimizes the integral functional

$$
\begin{equation*}
J=\int_{0}^{T}\left[\frac{1}{2} \rho(t)(x-h(t))^{2}+\frac{1}{2} M u^{2}\right] d t \tag{1.3}
\end{equation*}
$$

Here $\rho(t) \geqslant 0$ is a weight function, $M>0$ is some initially undetermined constant, to be chosen later using the acceleration constraint (1.2). In (1.3) the function $x(t)$ is related to the control $u(t)$ by the differential equations (1.1). The value of the $x$ coordinate at the initial and final instants $t=[0, T]$ are not assumed to be fixed, i.e. we have a problem with free end-points.
It is natural to specify the weight function $\rho(t)$ to be inversely proportional to the square of the meansquare error of the measurement of the coordinate $x(t)$. If the trajectory is reconstructed from discrete readings $h_{k}=h\left(t_{k}\right)$ at the instants $t_{k}, k=0, \ldots, N, t_{0}=0<t_{1}<\ldots<t_{N}=T$, made with mean square errors $\sigma_{k}$, then the function $\rho(t)$ is given by the formula

$$
\begin{equation*}
\rho(t)=\rho_{k} \delta\left(t-t_{k}\right), \quad \rho_{k}=1 / \sigma_{k}^{2} \tag{1.4}
\end{equation*}
$$

where $\delta(\cdot)$ is the Dirac $\delta$-function. We remark that to reconstruct a multidimensional trajectory the above problem must be solved independently for each spatial variable.

## 2. THE OPTIMAL CONTROL PROBLEM

To determine the optimal trajectories $x(t)$ and accelerations $u(t)$ which minimize functional (1.3), we use Pontryagin's maximum principle. In our case the Hamiltonian function is

$$
\begin{equation*}
H=\psi_{1} y+\psi_{2} u+1 / 2 \psi_{0}\left[\rho(t)(x-h(t))^{2}+M u^{2}\right] \tag{2.1}
\end{equation*}
$$

From this we obtain equations for the adjoint variables

$$
\begin{equation*}
\dot{\psi}_{1}=-\partial H / \partial x=-\psi_{0} \rho(t)(x-h(t)), \quad \dot{\psi}_{2}=-\partial H / \partial y=-\psi_{1}, \quad \dot{\psi}_{0}=0 \tag{2.2}
\end{equation*}
$$

Because both ends of the trajectory are free, the transversality conditions have the form

$$
\begin{equation*}
\psi_{i}(0)=\psi_{i}(T)=0, \quad i=1,2, \quad \psi_{0} \equiv-1 \tag{2.3}
\end{equation*}
$$

When $\psi_{0}=-1$ the function $H$ defined by (2.2) reaches its maximum in $u$ at the critical point, i.e. when $\partial H / \partial u=\psi_{2}-M u=0$.

From this we obtain

$$
\begin{equation*}
u=\psi_{2} / M \tag{2.4}
\end{equation*}
$$

Substituting (2.4) into system (1.1) and using (2.2), we find that the optimal trajectory is given by the set of equations

$$
\begin{equation*}
\dot{x}=y, \quad \dot{y}=\psi_{2} / M, \quad \Psi_{2}=-\psi_{1}, \quad \dot{\Psi}_{1}=\rho(t)(x-h(t)) \tag{2.5}
\end{equation*}
$$

with boundary conditions (2.3). After solving this boundary-value problem the optimal control is found from expression (2.4).

## 3. DISCRETE MEASUREMENTS

We will consider in more detail the discrete-measurement case which is important for practical applications. System (2.5) has an analytic solution when the function $\rho(t)$ is given by expression (1.4). Indeed, in the interval $\left(t_{k}, t_{k+1}\right), k=0, \ldots, N-1$, i.e. between the instants when measurements are made, the right-hand side of the last equation in (2.5) vanishes, and system (2.5) is easily integrated

$$
\begin{align*}
& x=x_{k}^{+}+y_{k}^{+} \tau+\frac{\psi_{2 k}^{+}}{2 M} \tau^{2}-\frac{\psi_{1 k}^{+}}{6 M} \tau^{3}, \quad y=y_{k}^{+}+\frac{\psi_{2 k}^{+}}{M} \tau-\frac{\psi_{1 k}^{+}}{2 M} \tau^{2}  \tag{3.1}\\
& \psi_{2}=\psi_{2 k}^{+}-\frac{\psi_{1 k}^{+}}{M} \tau, \quad \psi_{1}=\psi_{1 k}^{+}
\end{align*}
$$

Here $\tau=t-t_{k}$, and $x_{k}^{+}, y_{k}^{+}, \Psi_{2 k}^{+}, \psi_{1 k}^{+}$are the values of the corresponding variables at the instant $t=$ $t_{k}+0$ immediately after the $k$ th measurement. From (3.1) we conclude that in the interval ( $t_{k}, t_{k+1}$ ) the optimal trajectory $x(t)$ is the cubic spline.

The trajectory transition at the instant $t_{k}$ is given by jump conditions which follow from (2.5) and (1.4)

$$
\begin{equation*}
x_{k}^{+}=x_{k}, \quad y_{k}^{+}=y_{k}, \quad \psi_{2 k}^{+}=\psi_{2 k}, \quad \Psi_{1 k}^{+}=\psi_{1 k}+\rho_{k}\left(x_{k}-h_{k}\right) \tag{3.2}
\end{equation*}
$$

Here $x_{k}, y_{k}, \Psi_{1 k}, \Psi_{2 k}$ are values of the corresponding variables at the instant $t=t_{k}-0$ just before the $k$ th measurement. Thus, at the instant of measurement, either the third time derivative of the coordinate or the first derivative of the control is discontinuous; the other variables remain continuous.

Combining (3.1) and (3.2), we obtain the transformation $\Phi_{k}: R^{4} \rightarrow R^{4}$ governing the values of all variables at the instant $t=t_{k+1}-0$ in terms of the values at the instant $t=t_{k}-0$

$$
\begin{equation*}
x_{k+1}=x_{k}+y_{k} \tau_{k}+\frac{\Psi_{2 k}}{2 M} \tau_{k}^{2}-\frac{\Psi_{1, k+1}}{6 M} \tau_{k}^{3}, \quad y_{k+1}=y_{k}+\frac{\Psi_{2 k}}{M} \tau_{k}-\frac{\Psi_{1, k+1}}{2 M} \tau_{k}^{2} \tag{3.3}
\end{equation*}
$$

$$
\psi_{2, k+1}=\psi_{2 k}-\frac{\psi_{1, k+1}}{M} \tau_{k}, \quad \Psi_{1, k+1}=\psi_{1 k}+\rho_{k}\left(x_{k}-h_{k}\right)
$$

Here and below $\tau_{k}=t_{k+1}-t_{k}, k=0, \ldots, N-1$. We rewrite (3.3) in matrix form

$$
\begin{equation*}
z_{k+1}=A_{k} z_{k}+b_{k}, \quad k=0, \ldots, N-1 \tag{3.4}
\end{equation*}
$$

Here $z_{s}=\left(x_{s}, y_{s} \Psi_{2 s}, \Psi_{1 s}\right)^{T}, s=0, \ldots, N$, the symbol $T$ denotes transposition and we have introduced the notation

$$
A_{k}=\left\|\begin{array}{cccc}
1-\frac{\tau_{k}^{3}}{6 M} \rho_{k} & \tau_{k} & \frac{\tau_{k}^{2}}{2 M} & \frac{\tau_{k}^{3}}{6 M}  \tag{3.5}\\
-\frac{\tau_{k}^{2}}{2 M} \rho_{k} & 1 & \frac{\tau_{k}}{M} & \frac{\tau_{k}^{2}}{2 M} \\
-\tau_{k} \rho_{k} & 0 & 1 & -\tau_{k} \\
\rho_{k} & 0 & 0 & 1
\end{array}\right\|, b_{k}=\left\|\frac{\tau_{k}^{3}}{6 M}\right\| \frac{\tau_{k}^{2}}{2 M} \| \tau_{k} h_{k}
$$

Similarly, we introduce the mapping $\Phi_{k}^{\delta}: R^{4} \rightarrow R^{4}$ corresponding to the transformation (3.2) and write it in matrix form

$$
\begin{align*}
& z_{k}^{+}=A_{k}^{\delta} z_{k}+b_{k}^{\delta}, \quad k=0, \ldots, N \\
& A_{k}^{\delta}=\left\|\begin{array}{llll}
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & 1 & 0 \\
\rho_{k} & 0 & 0 & 1
\end{array}\right\|, b_{k}^{\delta}=\left\|\begin{array}{c}
0 \\
0 \\
0 \\
-\rho_{k} h_{k}
\end{array}\right\| \tag{3.6}
\end{align*}
$$

In order to find the optimal trajectory and control it is sufficient to determine the values of all components of the vector $z_{0}$ at the instant $t_{0}-0$. The values of the coordinate and control at all instants $t=t_{k}-0$ and $t=t_{k}+0, k=0, \ldots, N$ are then reproduced using mappings (3.4) and (3.6), while the values at the intermediate instants $t \in\left(t_{k}, t_{k+1}\right)$ are determined using formulae (3.1).

It would appear that the value of $z_{0}$ can be determined by the following simple method. Find the map

$$
\begin{equation*}
\boldsymbol{\Phi}=\boldsymbol{\Phi}_{N}^{\delta} \circ \boldsymbol{\Phi}_{N-1} \circ \ldots \circ \boldsymbol{\Phi}_{0} \tag{3.7}
\end{equation*}
$$

turning the initial point $z_{0}$ into the final point $z_{N}$. To determine variables satisfying boundary conditions (2.3) it is necessary to solve the system of linear equations

$$
\begin{equation*}
z_{N}^{+}=\Phi\left(z_{0}\right) \tag{3.8}
\end{equation*}
$$

Here two components for each of the vectors $z_{N}^{+}$and $z_{0}$ are known: $\psi_{N 1}^{+}=\psi_{N 2}^{+}=\psi_{01}=\psi_{02}=0$. Thus the problem of finding $z_{0}$ reduces to the algebraic operation of matrix multiplication in the calculation of the explicit form of the map (3.7) and to subsequent solution of the linear system of equations (3.8).

However, this simplest method fails because of fundamental numerical difficulties.
The problem is that for all $k$ the matrix $A_{k}$ from (3.5) has two eigenvalues with moduli greater than unity. The eigenvalues $\lambda$ of the matrix $A_{k}$ are governed by the characteristic equation

$$
\begin{equation*}
P(\lambda .)=\lambda^{4}+(\kappa / 6-4) \lambda^{3}+(2 \kappa / 3+6) \lambda^{2}+(\kappa / 6-4) \lambda+1=0, \quad \kappa=\rho_{k} \tau_{k}^{3} / M>0 \tag{3.9}
\end{equation*}
$$

This equation is reciprocal, and its roots are positioned symmetrically with respect to the unit circle. They would only lie on the unit circle itself when $\kappa=0$, that is impossible (see (3.9)). Hence there are always two roots with modulus greater than unity. The presence of these eigenvalues leads to instability and to an exponential increase in the matrix coefficients obtained when computing mapping (3.7), and an exponential increase in the numerical errors also. All this makes calculation impossible even for fairly small $N$. We must therefore abandon the calculation of mapping (3.7) and use another method.

## 4. MATRIX PIVOTAL CONDENSATION

We shall use the matrix pivotal condensation method [1]. We rewrite the matrices and vectors that occur in transformation (3.4) in the following partitioned form, together with the transformation itself

$$
\begin{align*}
& z_{k}=\left\|\begin{array}{l}
\xi_{1 k} \\
\xi_{2 k}
\end{array}\right\|, \quad \xi_{1 k}=\left\|\begin{array}{l}
x_{k} \\
y_{k}
\end{array}\right\|, \quad \xi_{2 k}=\left\|\begin{array}{l}
\Psi_{2 k} \\
\Psi_{1 k}
\end{array}\right\|, \quad A_{k}=\left\|\begin{array}{ll}
A_{11}^{k} & A_{12}^{k} \\
A_{21}^{k} & A_{22}^{k}
\end{array}\right\| \\
& b_{k}=\left\|\begin{array}{l}
b_{1 k} \\
b_{2 k}
\end{array}\right\|, k=0, \ldots, N  \tag{4.1}\\
& \xi_{1, k+1}=A_{11}^{k} \xi_{1 k}+A_{12}^{k} \xi_{2 k}+b_{1 k}, \quad \xi_{2, k+1}=A_{21}^{k} \xi_{1 k}+A_{22}^{k} \xi_{2 k}+b_{2 k}
\end{align*}
$$

Here $A_{i j}$ and $b_{i k}$ are $(2 \times 2)$-matrices and 2-vectors, corresponding to vectors $\xi_{1 k}$ and $\xi_{2 k}$. We make the substitution

$$
\begin{equation*}
\xi_{2 k}=Q_{k} \xi_{1 k}+q_{k}, \quad k=0, \ldots, N \tag{4.2}
\end{equation*}
$$

in (4.1), where $Q_{k}$ and $q_{k}$ are an initially unknown ( $2 \times 2$ )-matrix and 2-vector. We obtain

$$
\begin{align*}
& \xi_{1, k+1}=\left(A_{11}^{k}+A_{12}^{k} Q_{k}\right) \xi_{1 k}+A_{12}^{k} q_{k}+b_{1 k} \\
& Q_{k+1} \xi_{1, k+1}+q_{k+1}=\left(A_{21}^{k}+A_{22}^{k} Q_{k}\right) \xi_{1 k}+A_{22}^{k} q_{k}+b_{2 k} \tag{4.3}
\end{align*}
$$

We now substitute $\xi_{1, k+1}$ from the first equation in (4.3) into the second, and equating the coefficients of the $\xi_{1 k}$ and the free terms on the left- and right-hand sides, we obtain the system

$$
\begin{align*}
& Q_{k+1}\left(A_{11}^{k}+A_{12}^{k} Q_{k}\right)=A_{21}^{k}+A_{22}^{k} Q_{k} \\
& Q_{k+1}\left(A_{12}^{k} q_{k}+b_{1 k}\right)+q_{k+1}=A_{22} q_{k}+b_{2 k}, \quad k=0, \ldots, N-1 \tag{4.4}
\end{align*}
$$

Note that if the matrices $Q_{k}, Q_{k+1}$ and vectors $q_{k}, q_{k+1}$ satisfy system (4.4), the equation obtained from (4.3) after eliminating $\xi_{1, k+1}$ is satisfied identically by $\xi_{1 k}$. Relations (4.4) are recurrence relations connecting the $Q_{k}, q_{k}$ with the $Q_{k+1}, q_{k+1}$.

We now choose values for $Q_{N}, q_{N}$ so that boundary condition (2.3) at the right end of the trajectory is satisfied identically by any $\xi_{1 N}$. We recall that according to (4.1) the values $\xi_{1 N}, \xi_{2 N}$ correspond to values of the coordinates and adjoint variables at the instant $t_{N}-0$. Hence, in order to satisfy the boundary condition $\xi_{2}\left(t_{N}+0\right) \equiv 0$, it is necessary to apply transformation (3.6) to the vectors $\xi_{1 N}, \xi_{2 N}$. We obtain

$$
\xi_{2}\left(t_{N}+0\right)=\left\|\begin{array}{ll}
0 & 0  \tag{4.5}\\
\rho_{N} & 0
\end{array}\right\| \xi_{1 N}+\xi_{2 N}-\left\|\begin{array}{c}
0 \\
\rho_{N} h_{N}
\end{array}\right\| \equiv 0
$$

We substitute the expression for $\xi_{2 N}$, given by the substitution (4.2), into (4.5), and in the resulting relation we equate the coefficients of $\xi_{1 N}$ and the free terms to zero. We obtain

$$
Q_{N}=\left\|\begin{array}{ll}
0 & 0  \tag{4.6}\\
-\rho_{N} & 0
\end{array}\right\|, \quad q_{N}=\left\|\begin{array}{c}
0 \\
\rho_{N} h_{N}
\end{array}\right\|
$$

From relations (4.4) we obtain recurrence formulae expressing $Q_{k}, q_{k}$ in terms of $Q_{k+1}, q_{k+1}$

$$
\begin{equation*}
Q_{k}=\left(A_{22}^{k}-Q_{k+1} A_{12}^{k}\right)^{-1}\left(Q_{k+1} A_{11}^{k}-A_{21}^{k}\right), \quad q_{k}=\left(A_{22}^{k}-Q_{k+1} A_{12}^{k}\right)^{-1}\left(Q_{k+1} b_{1 k}-b_{2 k}+q_{k+1}\right) \tag{4.7}
\end{equation*}
$$

The first equation in (4.7) gives the non-linear relation between the matrices $Q_{k}$ and $Q_{k+1}$, while the second gives the linear relation between the vectors $q_{k}$ and $q_{k+1}$.

Using (4.7) and initial conditions (4.6) we sequentially obtain the matrices $Q_{N-1}, \ldots, Q_{0}$ and vectors $q_{N-1}, \ldots, q_{0}$. Having computed $Q_{0}$ and $q_{0}$, and then, using substitution (4.2) and boundary condition (2.3) at the initial instant ( $\xi_{20} \equiv 0$ ), we obtain the initial point of the optimal trajectory

$$
\left\|\begin{array}{l}
x_{0}  \tag{4.8}\\
y_{0}
\end{array}\right\|=\xi_{10}=-Q_{0}^{-1} q_{0}
$$

The initial value $z_{0}=\left(\xi_{10} \xi_{20}\right)^{T}$ is thus determined by relation (4.8). The optimal trajectory is now easily reconstructed at the nodes $t_{k}$ using expressions (4.2) and the first expression in (4.3), and at the intermediate points using (3.10).

For a final solution of the problem it remains to choose the parameter $M$ so that the acceleration condition (1.2) is satisfied. The maximum acceleration $u_{\max }$ along the optimal trajectory is a function of $M$, and $u_{\max } \rightarrow \infty$ as $M \rightarrow 0$ and $u_{\max } \rightarrow 0$ as $M \rightarrow \infty$. Using a trial method it is easy to find numerically the value of $M$ for which $\left|u_{\max }\right|=D$.

Thus the solution of the problem of reconstructing the trajectory in terms of measurement data reduces to calculating the matrices $Q_{k}$ and vectors $q_{k}$ by recurrence formulae (4.7) followed by a calculation of the trajectories from formulae (4.2), (4.3) and (3.10), together with the selection of the parameter $M$. We note that the most onerous part of this procedure-the calculation of the matrices $Q_{k}$ using the first (non-linear) relation in (4.7)-does not depend on the results of measurements of $h_{k}$ occurring in the vectors $b_{k}$ in accordance with (3.6). Hence the matrices $Q_{k}$ can be calculated in advance of the measurements. When different sets of measurements are performed with the same errors, there is no need to repeat the calculation of the $Q_{k}$, which is performed only once.

The algorithm described for solving the trajectory reconstruction problem was implemented on a computer. Calculations showed the effectiveness and high accuracy of the method.

## 5. STABILITY OF THE PIVOTAL CONDENSATION METHOD

The pivotal condensation method described in Section 4 is similar to the matrix pivotal condensation method described in [1, pp. 106-111]. However, a check showed that the sufficient conditions for the well-posedness and stability of the algorithm that were obtained in [1] are not satisfied in our case. We will show that those sufficient conditions can be widened considerably in cases when the $A_{k}$ matrices do not depend on the number $k$, i.e. when $A_{k}=A$, where $A$ is a constant matrix.

We shall say that the pivotal condensation method is well-posed for given end conditions $Q_{N} q_{N}$ in reverse time, if for all $k=N-1, \ldots, 0$ the matrices $Q_{k}$ and vectors $q_{k}$ are uniquely defined. From formulae (4.7) it obviously follows that the condition

$$
\begin{equation*}
\operatorname{det}\left(A_{22}-Q_{k+1} A_{12}\right) \neq 0, \quad k=N-1, \ldots, 0 \tag{5.1}
\end{equation*}
$$

is necessary and sufficient for the well-posedness Below we shall omit the superscript $k$ in the matrices $\boldsymbol{A}_{i j}^{k}$.

We shall say that the pivotal condensation method is asymptotically stable for end condition $Q_{N}$ in reverse time if the first of the mappings (4.7) has an asymptotically stable fixed point $Q_{k+1}=Q_{k}=Q$. and the matrix $Q_{N}$ lies in the domain of attraction of this point, and the second mapping in (4.7) has an asymptotically stable fixed point $q_{k+1}=q_{k}=0$ if $b_{1 k}, b_{2 k}$ are set equal to zero. Moreover, we shall require the first mapping in (4.3) to have an asymptotically stable fixed point $\xi_{1, k+1}=\xi_{1 k}=0$ if $q_{k}, b_{1 k}$ are set equal to zero.

We note that the first two conditions guarantee the computational stability of the determination of the condensational sequences of matrices $Q_{k}$ and vectors $q_{k}$, and the last conditions guarantee the stability of the reconstruction of the trajectory.

In a similar way one can define the well-posedness and asymptotic stability of the pivotal condensation method in forward time.

Before obtaining sufficient conditions for the well-posedness and asymptotic stability of the pivotal condensation method we will first derive transformation formulae for the matrices $Q_{k}$ and vectors $q_{k}$ for a linear change of variables in phase space. We consider the mapping

$$
\begin{equation*}
\Phi_{k}(z)=A z+b_{k}: R^{4} \rightarrow R^{4}, \quad k=0, \ldots, N-1 \tag{5.2}
\end{equation*}
$$

and the associated mapping of the pivotal condensation method (4.4). We make the linear change of variables

$$
\begin{equation*}
z=C z^{\prime}, \quad \Phi_{k}=C \Phi_{k}^{\prime} \tag{5.3}
\end{equation*}
$$

defined by the non-degenerate matrix $C$. Here and below, variables without a prime are old, and primed variables are new. We rewrite (5.3) in partitioned form

$$
\begin{align*}
& \xi_{1}=C_{11} \xi_{1}^{\prime}+C_{12} \xi_{2}^{\prime},  \tag{5.4}\\
& \xi_{2}=C_{21} \xi_{1}^{\prime}+C_{22} \xi_{2}^{\prime}
\end{align*} \quad C=\left\|\begin{array}{ll}
C_{11} & C_{12} \\
C_{21} & C_{22}
\end{array}\right\|, \quad z=\binom{\xi_{1}}{\xi_{2}}
$$

With this change of variables the mapping (5.2) becomes

$$
\begin{equation*}
\Phi_{k}^{\prime}\left(z^{\prime}\right)=A^{\prime} z^{\prime}+b_{k}, \quad A^{\prime}=C^{-1} A C, \quad b_{k}^{\prime}=C^{-1} b_{k}, \quad k=0, \ldots, N-1 \tag{5.5}
\end{equation*}
$$

We now derive transformation formulae for $Q_{k}, q_{k}$. We shall require them to satisfy the relations

$$
\begin{equation*}
\xi_{2 k}^{\prime}=Q_{k}^{\prime} \xi_{1 k}^{\prime}+q_{k}^{\prime}, \quad k=0, \ldots, N \tag{5.6}
\end{equation*}
$$

which are similar to (4.2). Substituting (5.6) into (5.4), and then into the expressions obtained for $\xi_{1 k}$ and $\xi_{2 k}$ in (4.2), we have

$$
\begin{equation*}
Q_{k}\left(C_{11}+C_{12} Q_{k}^{\prime}\right) \xi_{1 k}^{\prime}+Q_{k} C_{12} q_{k}^{\prime}+q_{k}=\left(C_{21}+C_{22} Q_{k}^{\prime}\right) \xi_{1 k}^{\prime}+C_{22} q_{k}^{\prime} \tag{5.7}
\end{equation*}
$$

Further, since the $Q_{k}^{\prime}, q_{k}^{\prime}$ do not depend on the vector $\xi_{1 k}^{\prime}$ (in the same way that $Q_{k}, q_{k}$ do not depend on $\xi_{1 k}$ ), we equate the coefficients of $\xi_{1 k}$ and the free terms on the left- and right-hand sides of (5.7). As a result we obtain the required transition formulae

$$
\begin{equation*}
Q_{k}\left(C_{11}+C_{12} Q_{k}^{\prime}\right)-\left(C_{21}+C_{22} Q_{k}^{\prime}\right)=0, \quad q_{k}=\left(C_{22}-Q_{k} C_{12}\right) q_{k}^{\prime} \tag{5.8}
\end{equation*}
$$

As well as (5.8), we obtain formulae connecting $Q_{k+1}^{\prime}, q_{k+1}^{\prime}$ with $Q_{k}^{\prime}, q_{k}^{\prime}$, derived in the same way as (4.7) for the old variables

$$
\begin{align*}
& Q_{k}^{\prime}=\left(A_{22}^{\prime}-Q_{k+1}^{\prime} A_{12}^{\prime}\right)^{-1}\left(Q_{k+1}^{\prime} A_{11}^{\prime}-A_{21}^{\prime}\right)  \tag{5.9}\\
& q_{k}^{\prime}=\left(A_{22}^{\prime}-Q_{k+1}^{\prime} A_{12}^{\prime}\right)^{-1}\left(Q_{k+1}^{\prime} b_{1 k}-b_{2 k}+q_{k+1}\right)
\end{align*}
$$

We now assume that the change of variables (5.3) reduces matrix $A$ to Jordan form.
In the general situation the eigenvalues $\lambda_{1}, \ldots, \lambda_{4}$ of the matrix $A$ are different and the matrix $A^{\prime}$ is diagonal. Hence $A_{12}^{\prime}=A_{21}^{\prime}=0$. Equations (5.9) can therefore be rewritten in the form

$$
\begin{equation*}
Q_{k}^{\prime}=\left(A_{22}^{\prime}\right)^{-1} Q_{k+1}^{\prime} A_{11}^{\prime}, \quad a_{k}^{\prime}=\left(A_{22}^{\prime}\right)^{-1}\left(Q_{k+1}^{\prime} b_{1 k}^{\prime}-b_{2 k}^{\prime}+q_{k+1}^{\prime}\right) \tag{5.10}
\end{equation*}
$$

To fix our ideas we take

$$
A_{11}^{\prime}=\left\|\begin{array}{cc}
\lambda_{1} & 0 \\
0 & \lambda_{2}
\end{array}\right\|, \quad A_{22}^{\prime}=\left\|\begin{array}{cc}
\lambda_{3} & 0 \\
0 & \lambda_{4}
\end{array}\right\|
$$

The relation between the matrices $Q_{k}^{\prime}$ and $Q_{k+1}^{\prime}$ is now linear, unlike relations (4.7). It is obvious that the pivotal condensation method in the new variables is always well-posed for any starting conditions $Q_{N}^{\prime} q_{N}^{\prime}$ and any number of iterations $N$ if $\lambda_{3} \neq 0, \lambda_{4} \neq 0$. Moreover, the mapping determined by the first formula in (5.10) has a unique fixed point $Q *=0$ which is globally asymptotically stable when max $\left(\left|\lambda_{1}\right|\right.$, $\left.\left|\lambda_{2}\right|\right)<\min \left(\left|\lambda_{3}\right|,\left|\lambda_{4}\right|\right)$ and unstable if the opposite inequality is satisfied. The map determined by the second formula in (5.10) has, if one sets $b_{1 k}^{\prime}$ and $b_{k}^{\prime}$ equal to zero, a fixed point $q^{*}=0$ which is globally asymptotically stable when $\min \left(\left|\lambda_{3}\right|,\left|\lambda_{4}\right|\right)>1$. The reconstructed trajectory is then generated by formula (5.6) together with

$$
\begin{equation*}
\xi_{1, k+1}^{\prime}=A_{11}^{\prime} \xi_{1 k}^{\prime}+b_{1 k}^{\prime} \tag{5.11}
\end{equation*}
$$

which is obtained in the same way as the first formula in (4.3). If one equates $b_{1 k}^{\prime}$ to zero, mapping (5.11) has a fixed point $\xi_{*}=0$ which is globally asymptotically stable when $\max \left(\left|\lambda_{1}\right|,\left|\lambda_{2}\right|\right)<1$.

The following theorem has therefore been proved.
Theorem 1. Suppose that the eigenvalues of the matrix $A$ defining the mapping (5.2) are distinct, and two of them lie inside, and two outside, the unit circle. Then, in variables in which the matrix $A$ has diagonal form, and its eigenvalues are arranged in order of increasing modulus $\left(\left|\lambda_{1}\right| \leqslant\left|\lambda_{2}\right|<1<\right.$
$\left|\lambda_{3}\right| \leqslant\left|\lambda_{4}\right|$, the pivotal condensation method is well-posed and asymptotically stable in reverse time for all initial conditions.

We note that from formulae (5.10)-(5.11) one can similarly obtain sufficient conditions for wellposedness and asymptotic stability of the pivotal condensation method in direct time, in the form $\min \left(\left|\lambda_{1}\right|,\left|\lambda_{2}\right|\right)>1>\max \left(\left|\lambda_{3}\right|,\left|\lambda_{4}\right|\right)$.

Theorem 2. Let the conditions of Theorem 1 be satisfied. If the starting value of the matrix $Q_{N}$ is such that

$$
\begin{equation*}
\operatorname{det}\left\|C_{22}-Q_{N} C_{22}\right\| \neq 0, \quad \operatorname{det}\left\|C_{11}+C_{12} Q_{0}^{\prime}\right\| \neq 0, \quad \operatorname{det}\left\|C_{21}+C_{22} Q_{0}^{\prime}\right\| \neq 0 \tag{5.12}
\end{equation*}
$$

then for an arbitrary value of $q_{N}$ the pivotal condensation method is well-posed and asymptotically stable in reverse time for the original variables.

Proof. For specified values of $Q_{N}$ and $q_{n}$ the sequences $Q_{k}$ and $q_{k}$ can be constructed not only from formulae (4.7), but also by means of the change of variables (5.3) by formulae (5.8) and (5.10), as shown in the diagram

$$
\left.\begin{array}{rl}
Q_{N}, q_{N} \xrightarrow{(4.7)} Q_{N-1}, q_{N-1} & \rightarrow \stackrel{(4.7)}{\rightarrow} Q_{0}, q_{0} \\
(5.8) \downarrow \\
Q_{N}, q_{N}^{\prime} \rightarrow Q_{N-1}^{\prime}, q_{N-1}^{\prime} & \rightarrow  \tag{5.8}\\
(5.10)
\end{array} Q_{(5.10)}^{\prime}, q_{0}^{\prime}\right)
$$

From relation (5.8) with $k=N$ and $k=0$ we obtain the following conditions of solvability. The transition from $Q_{N}, q_{N}$ to $Q_{N}^{\prime}, q_{N}^{\prime}$ is defined in det $\left\|C_{22}-Q_{N} C_{12}\right\| \neq 0$, while the transition from $Q_{0}^{\prime}, q_{0}^{\prime}$ to $Q_{0}, q_{0}$ is defined if det $\left\|C_{11}-C_{12} Q^{\prime}\right\| \neq 0$. Because the matrix $Q_{0}^{-1}$ is necessary to determine the initial conditions $x_{0}, y_{0}, Q_{0}$ must be non-degenerate. It follows from (5.8) that the necessary and sufficient condition for this is $\operatorname{det}\left\|C_{11}+C_{12} Q_{0}^{\prime}\right\| \neq 0$, $\operatorname{det}\left\|C_{21}+C_{22} Q_{0}^{\prime}\right\| \neq 0$. Theorem 2 is proved.

We will now apply the results obtained to the matrix $A$ defined by expression (3.5) with $\tau_{k}=\tau, \rho_{k}=$ $\rho$. Because the characteristic equation (3.9) of $A$ is reciprocal, the eigenvalues of this matrix are distinct and arranged symmetrically with respect to the unit circle. Hence the conditions of Theorem 1 are satisfied. By Theorem 2 the pivotal condensation method is well-posed and asymptotically stable if inequalities (5.12) are satisfied for finite values of $Q_{N}, q_{N}$ from (4.6). The validity of these inequalities can always be established numerically for chosen values of the system parameters $\tau, \rho M$. We note that the matrix $Q_{0}^{\prime}$ can be assumed to be zero after a relatively small number of iterations $N$, because its coefficients are of order $O\left(\varepsilon^{N}\right)$ where $\varepsilon=\left|\lambda_{2}\right| /\left|\lambda_{3}\right|<1$. In particular, for the example considered below we have $N=40, \varepsilon<1 / 2$. Hence we can put $Q_{0}^{\prime}=0$ in (5.12) with a high degree of accuracy, and then verify the conditions

$$
\operatorname{det}\left\|C_{22}-Q_{N} C_{12}\right\| \neq 0, \quad \operatorname{det}\left\|C_{11}\right\| \neq 0, \quad \operatorname{det}\left\|C_{21}\right\| \neq 0
$$

This verification was performed numerically in the example to be considered.

## 6. EXAMPLE

We shall consider a model problem for the motion of a helicopter, which from an initial state of horizontal flight performs a manoeuvre to increase its height. We assume that the horizontal component of its velocity is constant both in magnitude and direction. The vertical motion is the result of a controlling acceleration. Measurements of the height $h$ are made at intervals of $\tau=1 \mathrm{~s}$. The manoeuvre takes 15 s and the total time of the motion is 25 s . Thus $N=25$. The quantity $\sigma$ is taken to be 2 m . The result of smoothing the trajectory, according to the algorithm described above, when the absolute magnitude of the controlling acceleration does not exceed $D=2 \mathrm{~m} / \mathrm{s}^{2}$ (see (1.2)) is shown in Fig. 1. Time is shown horizontally in seconds and the height vertically in metres. The small circles show the measured values, and the continuous curves shows the smoothed trajectory. To illustrate the dependence of the degree of smoothness of the trajectory on the magnitude of the permitted acceleration, the same original trajectory was smoothed with a maximum permissible acceleration of $30 \mathrm{~m} / \mathrm{s}^{2}$. The result is shown by the dashed line.
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Fig. 1.

## REFERENCE

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