# Optimal Target Criteria for Stabilization Policy: Technical Appendix [Complete] 

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## A Proof of Lemmas and Propositions

## A. 1 Proof of Proposition 1

Here we show that there exists a unique bounded solution to the system of equations consisting of the linearized structural equations (2.8)-(2.10) together with the linearized FOCs (2.16)-(2.17). If we adjoin to these equations the identities

$$
\begin{align*}
\tilde{y}_{t} & =\tilde{y}_{t},  \tag{A.1}\\
E_{t} \tilde{\varphi}_{t+1} & =E_{t} \tilde{\varphi}_{t+1}, \tag{A.2}
\end{align*}
$$

then the system consisting of (2.8), (2.16), and (A.1)-(A.2) can be rewritten in matrix form as

$$
\begin{equation*}
\bar{M} E_{t} d_{t+1}=\bar{N} d_{t}-\bar{N}_{s} \bar{s}_{t} \tag{A.3}
\end{equation*}
$$

where $d_{t}$ is the $2(m+n)$-dimensional vector

$$
d_{t} \equiv\left[\begin{array}{c}
\tilde{\varphi}_{t} \\
\tilde{y}_{t-1} \\
E_{t} \tilde{\varphi}_{t+1} \\
\tilde{y}_{t},
\end{array}\right]
$$

$\bar{s}_{t}$ is a vector of exogenous disturbances that includes the elements of both $\tilde{\xi}_{t}$ and $\tilde{\xi}_{t-1}$, and

$$
\bar{M} \equiv\left[\begin{array}{cc}
\bar{M}_{11} & \bar{M}_{12}  \tag{A.4}\\
I_{m+n} & 0
\end{array}\right], \quad \bar{N} \equiv\left[\begin{array}{cc}
-\beta^{-1} \bar{M}_{12}^{\prime} & 0 \\
0 & I_{m+n}
\end{array}\right]
$$

where

$$
\bar{M}_{11} \equiv\left[\begin{array}{cc}
0 & \bar{A} \\
\bar{A}^{\prime} & S
\end{array}\right]=\bar{M}_{11}^{\prime}, \quad \text { and } \quad \bar{M}_{12} \equiv\left[\begin{array}{cc}
0 & -\bar{I} \\
0 & \beta R^{\prime}
\end{array}\right]
$$

Here we use the fact that $S$ is symmetric to obtain $\bar{M}_{11}=\bar{M}_{11}^{\prime}$.
In addition to conditions (A.3), the process $\left\{d_{t}\right\}$ must satisfy (2.9) and (2.17), and thus

$$
\begin{equation*}
F_{d}\left[d_{t}-E_{t-1} d_{t}\right]=F_{s}\left[\bar{s}_{t}-E_{t-1} \bar{s}_{t}\right] \tag{A.5}
\end{equation*}
$$

for all $t>t_{0}$, where $F_{d}$ is the $(m+n) \times 2(m+n)$ matrix

$$
F_{d} \equiv\left[\begin{array}{cccc}
S_{2} & 0 & 0 & 0 \\
0 & 0 & 0 & \bar{I}_{1} \\
0 & I_{m} & 0 & 0
\end{array}\right]
$$

using the notation

$$
S_{2} \equiv\left[\begin{array}{ll}
0 & I_{n-k}
\end{array}\right]
$$

for the $(n-k) \times n$ matrix that selects the last $n-k$ elements of any $n$-vector. (The first $n-k$ rows correspond to conditions (2.17), the next $k$ rows correspond to conditions (2.9), and the final $m$ rows state that the elements of $\tilde{y}_{t-1}$ cannot be affected by surprises in period t.)

In period $t_{0}$, the process must satisfy (2.10) and hence

$$
\begin{equation*}
F_{d} d_{t_{0}}=f_{t_{0}} \tag{A.6}
\end{equation*}
$$

where $f_{t_{0}}$ is a vector of $m+n$ initial conditions

$$
f_{t_{0}} \equiv\left[\begin{array}{c}
\tilde{\Theta}_{t_{0}-1} \\
\bar{A}_{1} \tilde{y}_{t_{0}-1}+\bar{D}_{\tilde{\xi}_{t_{0}}}-\beta \bar{F}_{t_{0}} \\
\tilde{y}_{t_{0}-1}
\end{array}\right]
$$

all of which are either predetermined or exogenous.
The following lemmas establish useful properties of the matrix pencil $\bar{M}-\mu \bar{N}$.
Lemma 7 Given Assumptions 2(b) and 3, the matrix pencil $\bar{M}-\mu \bar{N}$ is regular; that is, its determinant is non-zero for at least some complex $\mu$.

Proof. The determinant of $\bar{M}-\mu \bar{N}$ can be expressed as follows:

$$
\begin{align*}
\operatorname{det}(\bar{M}-\mu \bar{N}) & =\operatorname{det}\left[\begin{array}{cc}
\bar{M}_{11}+\mu \beta^{-1} \bar{M}_{12}^{\prime} & \bar{M}_{12} \\
I & -\mu I
\end{array}\right] \\
& =\operatorname{det}(-\mu I) \cdot \operatorname{det}\left[\left(\bar{M}_{11}+\mu \beta^{-1} \bar{M}_{12}^{\prime}\right)-\bar{M}_{12}(-\mu I)^{-1}\right] \\
& =\mu(-1)^{n+m} \cdot \operatorname{det}\left[\bar{M}_{11}+\mu \beta^{-1} \bar{M}_{12}^{\prime}+\mu^{-1} \bar{M}_{12}\right] \\
& =\mu(-1)^{n+m} \cdot \operatorname{det}\left[\begin{array}{cc}
0 & \bar{A}-\mu^{-1} \bar{I} \\
\bar{A}^{\prime}-\mu \beta^{-1} \bar{I}^{\prime} & S+\mu^{-1} \beta R^{\prime}+\mu R
\end{array}\right] . \tag{A.7}
\end{align*}
$$

The matrix pencil $\bar{M}-\mu \bar{N}$ is regular provided that its determinant is non-zero for at least some complex $\mu$.

Suppose the determinant is instead zero for all $\mu$. This means that there must exist finite-order vector polynomials $(\varphi(\mu), y(\mu))$ such that

$$
\left[\begin{array}{cc}
0 & \bar{A}-\mu^{-1} \bar{I}  \tag{A.8}\\
\bar{A}^{\prime}-\mu \beta^{-1} \bar{I}^{\prime} & S+\mu^{-1} \beta R^{\prime}+\mu R
\end{array}\right]\left[\begin{array}{l}
\varphi(\mu) \\
y(\mu)
\end{array}\right]=0
$$

for all $\mu \neq 0$, and $(\varphi(\mu), y(\mu))$ are not both equal to zero for all $\mu$. In addition, the solution cannot involve $y(\mu)=0$. For if there exists a function $\varphi(\mu) \equiv \sum_{i=0}^{k} \varphi_{i} \mu^{i}$ satisfying (A.8) with $y(\mu)=0$, one must have $\left[\bar{A}^{\prime}-\mu \beta^{-1} \bar{I}^{\prime}\right] \varphi(\mu)=0$. But this would imply that the function $\varphi(\mu) \equiv \sum_{i=0}^{k} \varphi_{k-i} \mu^{i}$ must satisfy (2.13), violating Assumption 2(b). Hence we must have $y(\mu) \neq 0$. Writing $y(\mu) \equiv \sum_{i=0}^{\infty} y_{i} \mu^{i}$ (where all but a finite number of the $y_{i}$ are zero), the first line of (A.8) implies that the sequence $\left\{y_{i}\right\}$ satisfies the hypotheses of Assumption 3. The second line implies that

$$
y\left(\mu^{-1} \beta\right)^{\prime}\left[S+\mu^{-1} \beta R^{\prime}+\mu R\right] y(\mu)=0
$$

for all $\mu$. Writing this expression in the form $\sum_{j=-(k+1)}^{k+1} \gamma_{j} \mu^{j}$, where $k$ is the order of $y(\mu)$, it follows that we must have $\gamma_{j}=0$ for all $j$. In particular, we must have $\gamma_{0}=0$. But $\gamma_{0}$ is just the left-hand side of (2.19), so this violates Assumption 3. It follows that $\operatorname{det}(\bar{M}-\mu \bar{N})$ cannot be zero for all $\mu$.

It is then possible to factor the polynomial $\operatorname{det}[\lambda \bar{M}-\rho \bar{N}]$ as

$$
\begin{equation*}
\prod_{i=1}^{2(n+m)}\left(\alpha_{i} \lambda-\beta_{i} \rho\right) \tag{A.9}
\end{equation*}
$$

where for any $i$, the complex numbers $\alpha_{i}$ and $\beta_{i}$ are not both equal to zero. Let $s$ be the number of factors for which $\alpha_{i} \neq 0$ and $\left|\beta_{i} / \alpha_{i}\right|<1$. There must then be $2 m+2 n-s$ factors for which $\beta_{i} \neq 0$ and $\left|\alpha_{i} / \beta_{i}\right| \leq 1$.

This implies that the matrices $\bar{M}$ and $\bar{N}$ can be decomposed as stated in the following lemma.

Lemma 8 Given Assumptions 2(b) and 3, there must exist non-singular $2(m+n) \times 2(m+n)$ real matrices $\bar{U}, \bar{V}$ such that

$$
\bar{U} \bar{M} \bar{V}=\left[\begin{array}{cc}
I_{s} & 0  \tag{A.10}\\
0 & \Omega
\end{array}\right], \quad \bar{U} \bar{N} \bar{V}=\left[\begin{array}{cc}
\Lambda^{\prime} & 0 \\
0 & I_{2 n+2 m-s}
\end{array}\right]
$$

Here $\Omega$ is a $(2 m+2 n-s) \times(2 m+2 n-s)$ real matrix for which all eigenvalues have modulus less than or equal to 1 while $\Lambda$ is an $s \times s$ real matrix for which all eigenvalues have modulus less than 1.

Proof. Under Assumptions 2(b) and 3, Lemma 7 implies that $\bar{M}-\mu \bar{N}$ is a real regular matrix pencil of dimensions $(2 m+2 n) \times(2 m+2 n)$. It follows from Theorem 3 of Gantmacher (1959, Chap. 12), or its version for a real canonical form proved in Appendix D, that there exist real invertible matrices $\tilde{U}, \bar{V}$ of dimensions $(2 m+2 n) \times(2 m+2 n)$ such that

$$
\tilde{U} \bar{M} \bar{V}=\left[\begin{array}{cc}
I & 0  \tag{A.11}\\
0 & \tilde{G}
\end{array}\right], \quad \tilde{U} \bar{N} \bar{V}=\left[\begin{array}{cc}
\tilde{H} & 0 \\
0 & I
\end{array}\right]
$$

where $\tilde{G}$ is an invertible matrix of the real Jordan form and $\tilde{H}$ is a real nilpotent matrix of the Jordan form.

Let us factor the polynomial $\operatorname{det}[\lambda \bar{M}-\rho \bar{N}]$ as in (A.9) and let $\nu(0 \leq \nu \leq 2 m+2 n)$ be the number of factors $\left(\alpha_{i} \lambda-\beta_{i} \rho\right)$ for which the complex numbers $\beta_{i}=0$, while the numbers $\alpha_{i}$ are necessarily nonzero. (Note that since the eigenvalues of $\bar{M}-\mu \bar{N}$ are the quantities $\alpha_{i} / \beta_{i}$, these $\nu$ factors correspond to the $\nu$ "infinite" eigenvalues of $\bar{M}-\mu \bar{N}$.) The existence of a decomposition of the form (A.11) implies that the factors of the characteristic polynomial $\operatorname{det}[\lambda \bar{M}-\rho \bar{N}]$ in (A.9) are the same as those of

$$
\operatorname{det}[\lambda I-\rho \tilde{H}] \cdot \operatorname{det}[\lambda \tilde{G}-\rho I]
$$

Since $\tilde{H}$ is nilpotent, $\operatorname{det}[\lambda I-\rho \tilde{H}]$ must correspond to the $\nu$ factors for which the $\beta_{i}=0$ and $\alpha_{i} \neq 0$. The matrix pencil $\lambda I-\rho \tilde{H}$ is thus of dimensions $\nu \times \nu$. This implies that the matrix pencil $\lambda \tilde{G}-\rho I$ is of dimensions $(2 m+2 n-\nu) \times(2 m+2 n-\nu)$ and its determinant is the product of the $2 m+2 n-\nu$ factors $\left(\alpha_{i} \lambda-\beta_{i} \rho\right)$ for which the complex numbers $\beta_{i} \neq 0$.
(The matrix $\tilde{G}$ has thus $2 m+2 n-\nu$ eigenvalues denoted by $\alpha_{i} / \beta_{i}$, all finite). Among these factors, let there be $2 m+2 n-s$ of them (with $0 \leq s \leq \nu$ ) for which $\beta_{i} \neq 0$ and $\left|\alpha_{i} / \beta_{i}\right| \leq 1$, so that there are $s-\nu$ factors for which $\beta_{i} \neq 0$ and $\left|\alpha_{i} / \beta_{i}\right|>1$. The latter $s-\nu$ factors necessarily have $\alpha_{i} \neq 0$ and $\left|\beta_{i} / \alpha_{i}\right|<1$.

Recalling that $\tilde{G}$ is in real Jordan from, this implies that it is possible to partition it as

$$
\left[\begin{array}{cc}
\tilde{G}_{1} & 0 \\
0 & \tilde{G}_{2}
\end{array}\right]
$$

where $\tilde{G}_{1} \in \mathbb{R}^{(s-\nu) \times(s-\nu)}$ is a block-diagonal matrix with eigenvalues satisfying $\left|\alpha_{i} / \beta_{i}\right|>1$, $\beta_{i} \neq 0$, and $\tilde{G}_{2} \in \mathbb{R}^{(2 m+2 n-s) \times(2 m+2 n-s)}$ is a block-diagonal matrix with eigenvalues satisfying $\left|\alpha_{i} / \beta_{i}\right| \leq 1, \beta_{i} \neq 0$. Since all eigenvalues of $\tilde{G}_{1}$ are nonzero, the matrix $\tilde{G}_{1}$ is non-singular. Combining the $s-\nu$ factors associated with $\tilde{G}_{1}$ with the $\nu$ factors associated with the matrix pencil $\lambda I-\rho \tilde{H}$ constitutes $s$ factors for which $\alpha_{i} \neq 0$ and $\left|\beta_{i} / \alpha_{i}\right|<1$.

It follows that the $2(n+m) \times 2(n+m)$ real matrix

$$
\bar{U} \equiv\left[\begin{array}{ccc}
I_{\nu} & 0 & 0 \\
0 & \tilde{G}_{1}^{-1} & 0 \\
0 & 0 & I_{2 m+2 n-s}
\end{array}\right] \tilde{U}
$$

is non-singular and satisfies (A.10), where $\Omega \equiv \tilde{G}_{2}$ is a $(2 m+2 n-s) \times(2 m+2 n-s)$ blockdiagonal matrix in real Jordan form with blocks corresponding to the factors of (A.9) for which $\left|\alpha_{i} / \beta_{i}\right| \leq 1$. This implies that

$$
\begin{equation*}
\|\Omega\| \leq 1 \tag{A.12}
\end{equation*}
$$

The matrix $\Lambda^{\prime} \equiv\left[\begin{array}{cc}\tilde{H} & 0 \\ 0 & \tilde{G}_{1}^{-1}\end{array}\right]$ is a $s \times s$ block-diagonal matrix in real Jordan form with $\nu$ zero eigenvalues (i.e., the eigenvalues of $\tilde{H}$ corresponding to the $\nu$ factors $\left(\alpha_{i} \lambda-\beta_{i} \rho\right)$ for which $\beta_{i}=0$ ), and another $s-\nu$ eigenvalues corresponding to the roots $\beta_{i} \neq 0,\left|\beta_{i} / \alpha_{i}\right|<1$. Thus all $s$ eigenvalues of $\Lambda$ satisfy $\left|\beta_{i} / \alpha_{i}\right|<1$, so that

$$
\begin{equation*}
\|\Lambda\|<1 \tag{A.13}
\end{equation*}
$$

Because the inequality (A.13) is strict, there also exist values $\delta>1$ such that

$$
\begin{equation*}
\|\delta \Lambda\|<1 \tag{A.14}
\end{equation*}
$$

In what follows, we shall consider a value of $\delta>1$ that is small enough for both (1.13) and (A.14) to hold.

Now let the matrices $\bar{U}, \bar{V}$ be partitioned conformably with the partitions in (A.10):

$$
\left.\bar{U}=\left[\begin{array}{c}
\bar{U}_{1}  \tag{A.15}\\
\bar{U}_{2}
\end{array}\right]\right\} \begin{aligned}
& s \text { rows } \\
& 2 m+2 n-s \text { rows },
\end{aligned} \quad \bar{V}=\left[\begin{array}{cc}
\bar{V}_{1} & \bar{V}_{2}
\end{array}\right]
$$

where $\bar{V}_{1}$ and $\bar{V}_{2}$ are respectively $2(n+m) \times s$ and $2(n+m) \times(2 m+2 n-s)$ matrices. It follows from the non-singularity of $\bar{U}$ and $\bar{V}$ that the columns of $\left[\begin{array}{cc}\bar{U}_{1}^{\prime} & \bar{U}_{2}^{\prime}\end{array}\right]$ form a basis for $\mathbb{R}^{2(n+m)}$, as do the columns of $\left[\begin{array}{cc}\bar{V}_{1} & \bar{V}_{2}\end{array}\right]$. Hence we can represent $d_{t}$ as

$$
d_{t}=\left[\begin{array}{ll}
\bar{V}_{1} & \bar{V}_{2}
\end{array}\right]\left[\begin{array}{l}
\psi_{t}  \tag{A.16}\\
\phi_{t}
\end{array}\right],
$$

where $\psi_{t}$ is of dimension $s$ and $\phi_{t}$ is of dimension $2 m+2 n-s$. The vectors $\left(\psi_{t}, \phi_{t}\right)$ can be uniquely re-constructed from the vector $d_{t}$, and vice versa.

The decomposition (A.10) defines stable and unstable subspaces for the matrix pencil $\bar{M}-\mu \bar{N}$. In particular, for any $\delta \geq 1$, let us define the $\delta$-stable subspace $\mathcal{D}_{\delta}$ as the set of values $d_{t_{0}}$ for which there exists a deterministic sequence $\left\{d_{t}\right\}$ for $t \geq t_{0}$ consistent with this value of $d_{t_{0}}$, satisfying

$$
\begin{equation*}
\bar{M} d_{t+1}=\bar{N} d_{t} \tag{A.17}
\end{equation*}
$$

for all $t \geq t_{0}$, and such that

$$
\begin{equation*}
\lim _{t \rightarrow \infty} \delta^{t} d_{t}=0 \tag{A.18}
\end{equation*}
$$

(In the case that $\delta=1$, we shall call $\mathcal{D} \equiv \mathcal{D}_{1}$ simply the stable subspace.) We then have the following result regarding the dimension of this linear space.

Lemma 9 Given Assumptions 2(b) and 3, let $\mathcal{D}_{\delta}$ be the $\delta$-stable subspace of the matrix pencil $\bar{M}-\mu \bar{N}$ corresponding to a value of $\delta$ such that (A.14) holds. Then $\mathcal{D}_{\delta}$ is a linear space of dimension $s$, the dimension of the square matrix $\Lambda$ in (A.10).

Proof. Under Assumptions 2(b) and 3, Lemma 8 holds, and we can then rewrite (A.10) as:

$$
\begin{array}{ll}
\bar{U}_{1} \bar{M} \bar{V}_{1}=I_{s}, & \bar{U}_{1} \bar{M} \bar{V}_{2}=0 \\
\bar{U}_{2} \bar{M} \bar{V}_{1}=0, & \bar{U}_{2} \bar{M} \bar{V}_{2}=\Omega \tag{A.20}
\end{array}
$$

and

$$
\begin{array}{lll}
\bar{U}_{1} \bar{N} \bar{V}_{1}=\Lambda^{\prime}, & \bar{U}_{1} \bar{N} \bar{V}_{2}=0 \\
\bar{U}_{2} \bar{N} \bar{V}_{1}=0, & \bar{U}_{2} \bar{N} \bar{V}_{2}=I_{2 m+2 n-s} \tag{A.22}
\end{array}
$$

We observe from these orthogonality relations that the inverse transformations can be written as

$$
\bar{U}^{-1}=\left[\begin{array}{ll}
\bar{M} \bar{V}_{1} & \bar{N} \bar{V}_{2}
\end{array}\right], \quad \bar{V}^{-1}=\left[\begin{array}{c}
\bar{U}_{1} \bar{M}  \tag{A.23}\\
\bar{U}_{2} \bar{N}
\end{array}\right]
$$

(Here we use the fact that because $\bar{U}$ and $\bar{V}$ are non-singular, we know that unique inverses exist.)

We can then pre-multiply the equations in (A.10) by $\bar{U}^{-1}$, using (A.23), to obtain:

$$
\begin{align*}
\bar{M} \bar{V}_{2} & =\bar{N} \bar{V}_{2} \Omega  \tag{A.24}\\
\bar{M} \bar{V}_{1} \Lambda^{\prime} & =\bar{N} \bar{V}_{1} . \tag{A.25}
\end{align*}
$$

We can similarly post-multiply the equations in (A.10) by $\bar{V}^{-1}$, using (A.23), to obtain:

$$
\begin{align*}
\Lambda^{\prime} \bar{U}_{1} \bar{M} & =\bar{U}_{1} \bar{N}  \tag{A.26}\\
\bar{U}_{2} \bar{M} & =\Omega \bar{U}_{2} \bar{N} \tag{A.27}
\end{align*}
$$

Because $\bar{U}^{-1}$ and $\bar{V}^{-1}$ must be non-singular matrices, we observe from (A.23) that $\bar{M} \bar{V}_{1}$, $\bar{N} \bar{V}_{2}, \bar{M}^{\prime} \bar{U}_{1}^{\prime}$, and $\bar{N}^{\prime} \bar{U}_{2}^{\prime}$ must each be matrices of full rank.

Pre-multiplying (A.17) by $\bar{U}_{2}$ and using (A.27), we obtain

$$
\Omega \bar{U}_{2} \bar{N} d_{t+1}=\bar{U}_{2} \bar{N} d_{t}
$$

for each $t \geq t_{0}$. Then using (A.16) to substitute for $d_{t}$ on both sides of this equation, and using (A.22), we obtain

$$
\Omega \phi_{t+1}=\phi_{t}
$$

which in turn implies that

$$
\begin{equation*}
\delta^{t} \phi_{t}=\left(\delta^{-1} \Omega\right) \delta^{t+1} \phi_{t+1} \tag{A.28}
\end{equation*}
$$

for each $t \geq t_{0}$. Repeated application of (A.28) implies that

$$
\begin{equation*}
\delta^{t} \phi_{t}=\left(\delta^{-1} \Omega\right)^{k} \delta^{t+k} \phi_{t+k} \tag{А.29}
\end{equation*}
$$

for arbitrary $k \geq 1$. Then in the case of any sequence $\left\{d_{t}\right\}$ satisfying (A.18), (A.12) implies that the right-hand side of (A.29) converges to zero for large $k$. Hence we must have $\phi_{t}=0$ for all $t \geq t_{0}$ in the case of any such sequence. Thus $d_{t}$ must be a vector of the form $d_{t}=\bar{V}_{1} \psi_{t}$ for all $t$.

Pre-multiplying (A.17) by $\bar{U}_{1}$ and again using (A.16) to substitute for $d_{t}$, one can similarly show that

$$
\psi_{t+1}=\Lambda^{\prime} \psi_{t}
$$

for all $t \geq t_{0}$. Given a vector $\psi_{t_{0}}$, this law of motion can be solved for the complete sequence $\left\{\psi_{t}\right\}$ and hence for the implied sequence $\left\{d_{t}\right\}$. Since

$$
\delta^{t} \psi_{t}=\left(\delta \Lambda^{\prime}\right)^{t-t_{0}} \delta^{t_{0}} \psi_{t_{0}}
$$

for any $t$, it follows from (A.14) that (A.18) must be satisfied. Hence the $\delta$-stable subspace $\mathcal{D}_{\delta}$ consists of all vectors of the form $d_{t_{0}}=\bar{V}_{1} \psi_{t_{0}}$ for some vector $\psi_{t_{0}}$. Since $\bar{V}$ is invertible, this linear space must be of dimension $s$ (the number of columns of $\bar{V}_{1}$ ).

We turn now to a further characterization of the dimension $s$. Since $\bar{M}-\mu \bar{N}$ is a regular pencil, a pair $(\lambda, \rho)$ determines an eigenvalue $\mu$ of $\bar{M}-\mu \bar{N}$ if $\operatorname{det}[\lambda \bar{M}-\rho \bar{N}]=0$ and $\rho-\lambda \mu=0$. (In particular, a pair $(\lambda, \rho)$ determines an infinite eigenvalue of $\bar{M}-\mu \bar{N}$ if $\operatorname{det}[\lambda \bar{M}-\rho \bar{N}]=0$, and $\rho \neq 0, \lambda=0$.) Because of the symmetries in the elements of the matrices $\bar{M}$ and $\bar{N}$, the eigenvalues of the pencil $\bar{M}-\mu \bar{N}$ also satisfy the following symmetry. ${ }^{23}$

[^1]Lemma 10 Given Assumptions 2(b) and 3, the set of values of $\mu=\rho / \lambda$ for which $\operatorname{det}[\lambda \bar{M}-$ $\rho \bar{N}]=0$ is such that if $\mu$ belongs to the set, so do the numbers $\beta \mu^{-1}$, and the complex conjugates $\bar{\mu}$ and $\overline{\beta \mu^{-1}}$. In particular, if the equation holds for $\rho=0$ (and arbitrary $\lambda$ ), then it also holds for $\lambda=0$ (and arbitrary $\rho$ ).

Proof. Given Assumptions 2(b) and 3, Lemma 7 implies that the matrix pencil $\bar{M}-\mu \bar{N}$ is regular. Hence the matrix pencil $\bar{M}-\mu \hat{N}$ where $\hat{N} \equiv \beta^{1 / 2} \bar{N}$ is also regular. Let us define the $2(n+m) \times 2(n+m)$ matrix

$$
J \equiv\left[\begin{array}{cc}
0 & I_{n+m} \\
-I_{n+m} & 0
\end{array}\right]
$$

and observe that

$$
\bar{M}^{\prime} J \bar{M}=\hat{N}^{\prime} J \hat{N},
$$

so that the transposed matrix pencil $(\bar{M}-\mu \hat{N})^{\prime}$ is symplectic. It follows that the generalized eigenvalues of the transposed pencil $(\bar{M}-\mu \hat{N})^{\prime}$ are symmetric with respect to the unit circle (see Theorems 4 and 5 of Pappas, Laub and Sandell, 1980): if $\mu \in \mathbb{C}$ is a generalized eigenvalue of the real matrix pencil $(\bar{M}-\mu \hat{N})^{\prime}$, then so are $\mu^{-1}$ and the complex conjugates $\bar{\mu}, \overline{\mu^{-1}}$. In particular, if $\mu=0$ is an eigenvalue of $(\bar{M}-\mu \hat{N})^{\prime}$, so is $\mu=\infty$.

Since $\operatorname{det}[\bar{M}-\mu \hat{N}]=\operatorname{det}\left[\bar{M}^{\prime}-\mu \hat{N}^{\prime}\right]$ for all $\mu$, it follows that if $\mu \in \mathbb{C}$ is an eigenvalue of $(\bar{M}-\mu \hat{N})$, then so are $\mu^{-1}$ and the complex conjugates $\bar{\mu}, \overline{\mu^{-1}}$. Moreover, $\operatorname{det}[\lambda \bar{M}-\rho \hat{N}]=0$ if and only if $\operatorname{det}\left[\lambda \bar{M}-\beta^{1 / 2} \rho \bar{N}\right]=0$. Hence $\mu$ is a generalized eigenvalue of $(\bar{M}-\mu \bar{N})$ if and only if $\beta^{-1 / 2} \mu$ is a generalized eigenvalue of the transformed pencil $(\bar{M}-\mu \hat{N})$. It then follows that $\beta \mu^{-1}, \bar{\mu}$, and $\overline{\beta \mu^{-1}}$ must also be generalized eigenvalues of $(\bar{M}-\mu \bar{N})$.

Lemma 11 Given Assumptions 1(b), 2(b) and 3, the dimension of the square matrix $\Lambda$ in the decomposition (A.10) must be exactly $s=m+n$. Hence the matrix pencil $\bar{M}-\mu \bar{N}$ has exactly $m+n$ generalized eigenvalues satisfying $|\mu|<\beta$ and another $m+n$ generalized eigenvalues (some of which may be infinite) satisfying $|\mu|>1$, and the stable subspace $\mathcal{D}$ is of dimension $m+n$. The dimension of the square matrix $\Omega$ is also $m+n$, and this matrix satisfies

$$
\begin{equation*}
\|\Omega\|<\beta \tag{A.30}
\end{equation*}
$$

Proof. Assumption 1(b) implies that for any initial conditions close enough to consistency with the optimal steady state, there must exist a solution to the first-order conditions (for the deterministic case in which $\xi_{t}=\bar{\xi}$ at all times) in which (1.13) holds. It follows that for arbitrary initial conditions $f_{t_{0}}$, there must be a sequence $\left\{d_{t}\right\}$ satisfying the linearized FOCs (A.17) for all $t \geq t_{0}$, such that $d_{t_{0}}$ is consistent with (A.6), and such that $\left\{\delta^{t} d_{t}\right\}$ is a bounded sequence. ${ }^{24}$ It is then furthermore possible to choose a $\delta>1$ (possibly slightly smaller than the $\delta$ referred to in Assumption 1(b)) such that (A.18) is satisfied. For a small enough choice of $\delta>1$, (A.14) must hold as well. Hence there must exist $\delta>1$ for which (A.14) holds, and

[^2]such that for any initial conditions $f_{t_{0}}$, there exists a vector $d_{t_{0}}$ in the $\delta$-stable subspace $\mathcal{D}_{\delta}$ consistent with (A.6).

It follows from our characterization of the $\delta$-stable subspace in the proof of Lemma 9 that there must exist a vector $\psi_{t_{0}}$ such that

$$
\begin{equation*}
\left[F_{d} \bar{V}_{1}\right] \psi_{t_{0}}=f_{t_{0}} \tag{A.31}
\end{equation*}
$$

It is easily seen that any values for the $m+n$ elements of $f_{t_{0}}$ can be arranged through a suitable specification of the $n-k$ elements of $\tilde{\Theta}_{t_{0}-1}$, the $k$ elements of $\bar{F}_{t_{0}}$, and the $m$ elements of $\tilde{y}_{t_{0}-1}$. Hence the right-hand side of (A.31) can be any element of $\mathbf{R}^{m+n}$. Then in order for a solution to exist for arbitrary initial conditions, it is necessary that

$$
\begin{equation*}
\operatorname{rank} F_{d} \bar{V}_{1}=m+n \tag{A.32}
\end{equation*}
$$

This requires that $s \geq m+n$.
We further note that the decomposition (A.10) implies that the generalized eigenvalues of the pencil $\bar{M}-\mu \bar{N}$ consist of the $2 m+2 n-s$ eigenvalues $\mu_{i}$ of the matrix $\Omega$ and the reciprocals of the $s$ eigenvalues $\lambda_{j}$ of the matrix $\Lambda$. Lemma 10 implies that for each eigenvalue $\lambda_{j}$ of $\Lambda, \beta \lambda_{j}$ must also be a generalized eigenvalue of the pencil $\bar{M}-\mu \bar{N}$; and since $\left|\lambda_{j}\right|<1$, this must be a generalized eigenvalue with modulus less than $\beta$, and therefore an eigenvalue of $\Omega$ rather than the reciprocal of any eigenvalue of $\Lambda$. Hence for each eigenvalue $\lambda_{j}$ of $\Lambda$, $\beta \lambda_{j}$ must be an eigenvalue of $\Omega$. This requires that $\Omega$ be of at least the dimension of $\Lambda$, and hence that $s \leq m+n$. Therefore $s=m+n$ exactly. The matrix $\Omega$ is of dimension $m+n$, and its eigenvalues all satisfy $|\mu|<\beta$, which implies (A.30).

Finally, it follows that $\left[F_{d} \bar{V}_{1}\right]$ must be a non-singular square matrix, so that (A.31) can be solved for $\psi_{t_{0}}$ for any specification of the initial conditions $f_{t_{0}}$. Since the largest eigenvalue of $\Lambda$ must have a modulus strictly less than 1 , any initial condition of the form $d_{t_{0}}=\bar{V}_{1} \psi_{t_{0}}$ gives rise to a sequence $\left\{d_{t}\right\}$ satisfying (A.18) for $\delta=1$. Hence this linear space of dimension $m+n$ corresponds to the stable subspace.

In the proof of Lemma 11, it has already been established that for any initial conditions $f_{t_{0}}$, there exists a deterministic solution $\left\{d_{t}\right\}$ to the linearized FOCs that converges exponentially to the steady state for large $t$. This result can then be directly extended to the case of bounded fluctuations in the exogenous disturbances $\left\{\tilde{\xi}_{t}\right\}$, yielding the result stated in the proposition.

Given a bounded stochastic process $\left\{\tilde{\xi}_{t}\right\}$ for the exogenous disturbances and a vector $f_{t_{0}}$ of initial conditions, we are interested in stochastic processes $\left\{d_{t}\right\}$ such that (i) $\left\{d_{t}\right\}$ is bounded; (ii) (A.3) is satisfied for all $t \geq t_{0}$; (iii) (A.5) is satisfied for all $t>t_{0}$; and (iv) $d_{t_{0}}$ satisfies (A.6). Pre-multiplying (A.3) by $\bar{U}_{2}$, we can show as in the proof of Lemma 9 that

$$
\Omega \bar{U}_{2} \bar{N} E_{t} d_{t+1}=\bar{U}_{2} \bar{N} d_{t}-\bar{U}_{2} \bar{N}_{s} \bar{s}_{t}
$$

or equivalently that

$$
E_{t}\left[\left(I-\Omega L^{-1}\right) \bar{U}_{2} \bar{N} d_{t}\right]=\bar{U}_{2} \bar{N}_{s} \bar{s}_{t} .
$$

Using (A.16) to substitute for $d_{t}$, this can alternatively be written

$$
\begin{equation*}
E_{t}\left[\left(I-\Omega L^{-1}\right) \phi_{t}\right]=\bar{U}_{2} \bar{N}_{s} \bar{s}_{t} . \tag{А.33}
\end{equation*}
$$

Because of (A.30), the operator $I-\Omega L^{-1}$ is invertible on the linear space of bounded processes $\left\{\phi_{t}\right\}$, so that for any disturbance process such that $\|\bar{s}\|<\infty$, (A.33) has a unique solution such that $\|\phi\|<\infty$, given by

$$
\begin{equation*}
\phi_{t}=E_{t}\left[\left(I-\Omega L^{-1}\right)^{-1} \bar{U}_{2} \bar{N}_{s} \bar{s}_{t}\right] . \tag{А.34}
\end{equation*}
$$

Similarly, pre-multiplying (A.3) by $\bar{U}_{1}$ and using (A.16) to substitute for $d_{t}$ yields

$$
\begin{equation*}
E_{t} \psi_{t+1}=\Lambda^{\prime} \psi_{t}-\bar{U}_{1} \bar{N}_{s} \bar{s}_{t} . \tag{A.35}
\end{equation*}
$$

Using (A.16) to substitute for $d_{t}$ in (A.5), and shifting the time index by one period, yields

$$
F_{d} \bar{V}_{1}\left[\psi_{t+1}-E_{t} \psi_{t+1}\right]=F_{s}\left[\bar{s}_{t}-E_{t-1} \bar{s}_{t}\right]-F_{d} \bar{V}_{2}\left[\phi_{t+1}-E_{t} \phi_{t+1}\right]
$$

for each $t \geq t_{0}$. Since $F_{d} \bar{V}_{1}$ is an invertible square matrix (as shown in the proof of Lemma 11 ), this can be solved uniquely for $\psi_{t+1}$. Substituting expression (A.35) for the conditional expectation $E_{t} \psi_{t+1}$ in this equation, and the solution (A.34) for both $\phi_{t+1}$ and its conditional expectation, we obtain a law of motion of the form

$$
\begin{equation*}
\psi_{t+1}=\Lambda^{\prime} \psi_{t}+r_{t+1} \tag{A.36}
\end{equation*}
$$

for all $t \geq t_{0}$, where $\left\{r_{t}\right\}$ is a process satisfying $\|r\|<\infty$ that has been uniquely determined as a function of the evolution of the exogenous disturbances.

Finally, using (A.16) to substitute for $d_{t_{0}}$ in (A.6) we obtain

$$
F_{d} \bar{V}_{1} \psi_{t_{0}}=f_{t_{0}}-F_{d} \bar{V}_{2} \phi_{t_{0}} .
$$

Using the solution (A.34) to substitute for $\phi_{t_{0}}$ in this equation, the invertibility of $F_{d} \bar{V}_{1}$ implies that this equation has a unique solution for $\psi_{t_{0}}$ for any specification of the initial conditions $f_{t_{0}}$ and the process for the exogenous disturbances. Given this initial condition for $\psi_{t_{0}}$, the law of motion (A.36) can then be integrated forward, yielding a unique solution for the evolution of $\left\{\psi_{t}\right\}$ for all $t \geq t_{0}$. It follows from (A.14) and the fact that $\|r\|<\infty$ that this solution will satisfy $\|\psi\|<\infty$. Our solutions for the processes $\left\{\phi_{t}, \psi_{t}\right\}$ then imply a unique solution for the process $\left\{d_{t}\right\}$, using (A.16), and the bounds satisfied by the two solutions imply that $\|d\|<\infty$ as well. Hence there is a unique solution satisfying this bound. QED.

## A. 2 Proof of Lemma 1

There is no vector $\hat{\varphi}_{2}(\mu) \neq 0$ such that $\hat{\varphi}_{2}(\mu)^{\prime}\left[J_{2}-\mu B_{2}\right]=0$ for all $\mu$. For if there did, the vector

$$
\varphi(\mu)=P^{-1}\left[\begin{array}{c}
0 \\
\hat{\varphi}_{2}(\mu)
\end{array}\right] \neq 0
$$

would satisfy (2.13), and this would violate Assumption 2. This implies that the pencil $J_{2}-\mu B_{2}$ must be of rank $q$, which (since it is a square pencil of dimension $q \times q$ ) implies that it is a regular pencil.

It follows from Theorem 3 of Gantmacher (1959, Chap. 12), or its version for a real canonical form proved in Appendix D, that $J_{2}-\mu B_{2}$ can be reduced to a strictly equivalent pencil of the form

$$
\left[\begin{array}{cc}
I_{l}-\mu \hat{G}^{\prime} & 0  \tag{А.37}\\
0 & \hat{H}^{\prime}-\mu I_{q-l}
\end{array}\right]
$$

where $\hat{G}^{\prime} \in \mathbb{R}^{l \times l}(0 \leq l \leq q)$ is a nilpotent matrix of the Jordan form (i.e., with ones on the first super diagonal and zeros everywhere else), and $\hat{H}^{\prime} \in \mathbb{R}^{(q-l) \times(q-l)}$ is a block-diagonal matrix of the real Jordan form. We can without loss of generality arrange the Jordan blocks of $\hat{H}^{\prime}$ as

$$
\hat{H}^{\prime}=\left[\begin{array}{cc}
\hat{H}_{11}^{\prime} & 0 \\
0 & H^{\prime}
\end{array}\right]
$$

where the invertible matrix $\hat{H}_{11}$ contains the eigenvalues with modulus greater than or equal to $\beta$, and $H$ contains only eigenvalues with modulus less than $\beta$. Premultiplying the pencil (A.37) by the invertible block-diagonal matrix

$$
\left[\begin{array}{ccc}
I_{l} & 0 & 0 \\
0 & \left(\hat{H}_{11}^{\prime}\right)^{-1} & 0 \\
0 & 0 & I_{q-k_{2}}
\end{array}\right]
$$

yields a strictly equivalent matrix pencil of the form (2.29), where all eigenvalues of

$$
G^{\prime}=\left[\begin{array}{cc}
\hat{G}^{\prime} & 0 \\
0 & \left(\hat{H}_{11}^{\prime}\right)^{-1}
\end{array}\right] .
$$

have modulus less than or equal to $\beta^{-1}$ and all eigenvalues of $H$ have modulus less than $\beta$. It remains only to determine the dimensions of $G$ and $H$.

The existence of a decomposition of the form (2.29) implies that the factors of the characteristic polynomial $P(\lambda, \mu)$ defined in Assumption 4 are the same as those of

$$
\operatorname{det}\left[\lambda I-\mu G^{\prime}\right] \cdot \operatorname{det}\left[\lambda H^{\prime}-\mu I\right]
$$

This implies the existence of a factorization of the form (2.22), where the $\left\{\gamma_{i}\right\}$ are the eigenvalues of $G$ and the $\left\{\eta_{j}\right\}$ are the eigenvalues of $H$. It follows that $G$ must be of dimension $k_{2} \times k_{2}$ and $H$ must be of dimension $\left(q-k_{2}\right) \times\left(q-k_{2}\right)$. Finally, since by Assumption 4(b), $\left|\gamma_{i}\right|<1$ for all $i$, all eigenvalues of $G$ must have modulus strictly smaller than 1 .

## A. 3 Proof of Proposition 2

Premultiplying (2.24) by $N_{2}^{\prime}$ yields the pair of equations

$$
\begin{align*}
E_{t} x_{1, t+1} & =G^{\prime} x_{1, t}+N_{21}^{\prime} \Gamma_{2} \tilde{\xi}_{t}  \tag{A.38}\\
H^{\prime} E_{t} x_{2, t+1} & =x_{2, t}+N_{22}^{\prime} \Gamma_{2} \tilde{\xi}_{t} \tag{A.39}
\end{align*}
$$

using the decomposition (2.29), defining

$$
x_{t} \equiv\left[\begin{array}{l}
x_{1 t}  \tag{A.40}\\
x_{2 t}
\end{array}\right] \equiv\left(R_{2}^{-1}\right)^{\prime} y_{2 t}^{*},
$$

and partitioning $N_{2}^{\prime} \equiv\left[\begin{array}{c}N_{21}^{\prime} \\ N_{22}^{\prime}\end{array}\right]$ conformably with the partition of $x_{t}$.
Because all eigenvalues of $H^{\prime}$ have modulus less than $\beta$, and $\left\{\tilde{\xi}_{t}\right\}$ is bounded for all dates $t \geq t_{0}-1$, there is a unique process $\left\{x_{2 t}\right\}$ consistent with (A.39) and such that $\left\|x_{2}\right\|<\infty$, namely

$$
\begin{equation*}
x_{2 t}=-\sum_{j=0}^{\infty}\left(H^{\prime}\right)^{j} N_{22}^{\prime} \Gamma_{2} E_{t} \tilde{\xi}_{t+j} \tag{A.41}
\end{equation*}
$$

Then in any period $t$, given the past, current and expected future values of the exogenous disturbance process, and given the lagged expectations of $J_{2} E_{t-1} y_{2 t}^{*}$, equation (2.26) determines the value of $X_{1}^{\prime} J_{2} y_{2, t}^{*}$, while equation (A.41) determines the value of $x_{2 t}$ and hence of $H^{\prime} x_{2 t}$. We then have a system of equations of the form

$$
\left.\left[\begin{array}{cc} 
& X_{1}^{\prime} \\
{[0} & I_{q-k_{2}}
\end{array}\right] N_{2}^{\prime}\right] J_{2} y_{2, t}^{*}=\left[\begin{array}{c}
X_{1}^{\prime} J_{2} y_{2, t}^{*} \\
H^{\prime} x_{2 t}
\end{array}\right]
$$

to solve for $J_{2} y_{2, t}^{*}$, where all elements of the matrix on the right-hand side have been computed. Since the matrix on the left-hand side is invertible by Assumption 5, this system has a unique solution for $J_{2} y_{2, t}^{*}$. Using this solution, we can in turn solve for

$$
x_{1, t}=\left[\begin{array}{ll}
I & 0
\end{array}\right] N_{2}^{\prime} J_{2} y_{2, t}^{*} .
$$

Combining this solution for $x_{1 t}$ with (A.41), we have a unique solution for the entire vector $x_{t}$, given values of $J_{2} E_{t-1} y_{2, t}^{*}$ and the evolution of the exogenous disturbances. And given the value of $x_{t}$ in any period, equations (A.38) and (A.39) uniquely determine the values of $E_{t}\left[x_{1, t+1}\right]$ and $E_{t}\left[H^{\prime} x_{2, t+1}\right]$ respectively. This allows us to uniquely determine

$$
J_{2} E_{t} y_{2, t+1}^{*}=N_{2}^{-1}\left[\begin{array}{c}
E_{t} x_{1, t+1} \\
E_{t}\left[H^{\prime} x_{2, t+1}\right]
\end{array}\right]
$$

Thus starting from given initial conditions $J_{2} E_{t_{0}-1} y_{2, t_{0}}^{*}$, we can uniquely solve for $x_{t_{0}}$, use this to uniquely solve for $J_{2} E_{t_{0}} y_{2, t_{0}+1}^{*}$, use this to uniquely solve for $x_{t_{0}+1}$, and so on recursively, eventually obtaining a unique solution for the entire process $\left\{x_{t}\right\}$, and hence a unique solution for the entire process $\left\{y_{2 t}^{*}\right\}$, using the relation $y_{2 t}^{*}=R_{2}^{\prime} x_{t}$.

This solution $\left\{y_{2 t}^{*}\right\}$ is the only solution such that $\left\|y_{2}^{*}\right\|<\infty$, if any solution exists. But one easily verifies that it is indeed such a solution. By construction, (2.26) is satisfied each period, and also both (A.38) and (A.39), which suffice to imply that (2.24) is satisfied each period. Moreover, the fact that all eigenvalues of $G$ have modulus less than 1 implies that the process $\left\{x_{t}\right\}$ constructed in this way satisfies $\|x\|<\infty$, so that the associated process $\left\{y_{2 t}^{*}\right\}$ satisfies $\left\|y_{2}^{*}\right\|<\infty$. Hence all conditions for a solution are satisfied.

## A. 4 Proof of Lemma 2

Using the definition of $q_{t}$ and the fact that the matrix $W$ is symmetric, we may rewrite the objective function (3.3) as

$$
\pi\left(y_{t} ; \xi_{t}\right)=-\frac{1}{2}\left[y_{t}^{\prime} w W w^{\prime} y_{t}-2 q^{*}\left(\xi_{t}\right)^{\prime} W w^{\prime} y_{t}+q^{*}\left(\xi_{t}\right)^{\prime} W q^{*}\left(\xi_{t}\right)\right]
$$

so that

$$
\begin{align*}
D_{1} \pi\left(y_{t} ; \xi_{t}\right) & =-\left[y_{t}^{\prime} w-q^{*}\left(\xi_{t}\right)^{\prime}\right] W w^{\prime}  \tag{A.42}\\
D_{1}\left[\left(D_{1} \pi\left(y_{t} ; \xi_{t}\right)\right)^{\prime}\right] & =-w W w^{\prime} \\
D_{2}\left[\left(D_{1} \pi\left(y_{t} ; \xi_{t}\right)\right)^{\prime}\right] & =w W\left[D q^{*}\left(\xi_{t}\right)\right] .
\end{align*}
$$

The fact that the targets $q^{*}\left(\xi_{t}\right)$ are achievable implies that in steady state, $D_{1} \pi(\bar{y}, \bar{\xi})=0$, using (3.4) and (A.42). It then follows from (1.12) that

$$
\bar{\varphi}^{\prime}(\bar{I}-\beta \bar{A})=0
$$

Assumption 2(b) then implies that $\bar{\varphi}=0$, so that $S, R$ and $B(L)$ reduce to

$$
S=-w W w^{\prime}, \quad R=0, \quad B(L)=w W\left[D q^{*}(\bar{\xi})\right] \cdot L
$$

The target variables and target values are then given by

$$
\begin{aligned}
\tau_{t} & =w W w^{\prime} \tilde{y}_{t}=w W\left(q_{t}-\bar{q}\right) \\
\tau_{t}^{*} & =w W\left[D q^{*}(\bar{\xi})\right] \cdot \tilde{\xi}_{t}
\end{aligned}
$$

where $\bar{q} \equiv w^{\prime} \bar{y}$. Performing a first-order approximation to $q^{*}\left(\xi_{t}\right)$, we obtain

$$
q_{t}^{*} \equiv q^{*}\left(\xi_{t}\right)=q^{*}(\bar{\xi})+\left[D q^{*}(\bar{\xi})\right] \cdot \tilde{\xi}_{t}+\mathcal{O}\left(\epsilon^{2}\right)
$$

so that, using (3.4), we have

$$
\left[D q^{*}(\bar{\xi})\right] \cdot \tilde{\xi}_{t}=q_{t}^{*}-\bar{q}+\mathcal{O}\left(\epsilon^{2}\right)
$$

Using this, we can express the "target gaps" $\tau_{t}-\tau_{t}^{*}$ as

$$
\tau_{t}-\tau_{t}^{*}=w W\left(q_{t}-\bar{q}\right)+w W\left(q_{t}^{*}-\bar{q}\right)+\mathcal{O}\left(\epsilon^{2}\right)=w W\left(q_{t}-q_{t}^{*}\right)+\mathcal{O}\left(\epsilon^{2}\right)
$$

## A. 5 Proof of Lemma 3

The fact that the pencil $\bar{A}-\mu \bar{I}$ is of rank $n<m$ implies that the columns are linearly dependent, i.e., that there exists $y(\mu)$ such that

$$
\begin{equation*}
[\bar{A}-\mu \bar{I}] y(\mu)=0 \tag{A.43}
\end{equation*}
$$

for all $\mu$, though by Assumption 2, $y(\mu)$ is of order greater than zero. Let $\epsilon_{1} \geq 1$ be the minimal order of solution $y(\mu)$ that exists to (A.43). (A solution of finite order $\epsilon_{1}$ necessarily exists.) Then Theorem 4 of Gantmacher (1959, chap. 12) implies that the pencil $\bar{A}-\mu \bar{I}$ is strictly equivalent to a pencil of the form

$$
\left[\begin{array}{cc}
M_{1}(\mu)^{\prime} & 0 \\
0 & B_{2}^{(1)}-\mu J_{2}^{(1)}
\end{array}\right]
$$

where $B_{2}^{(1)}-\mu J_{2}^{(1)}$ is a pencil for which the equation

$$
\begin{equation*}
\left[B_{2}^{(1)}-\mu J_{2}^{(1)}\right] \hat{y}^{(1)}(\mu)=0 \tag{A.44}
\end{equation*}
$$

has no solution of order less than $\epsilon_{1}$.
If (A.44) nonetheless has a nonzero solution of minimal order $\epsilon_{2} \geq \epsilon_{1}$, then Theorem 4 of Gantmacher can be applied again to the pencil $B_{2}^{(1)}-\mu J_{2}^{(1)}$. Proceeding in this way, one eventually transforms the pencil $\bar{A}-\mu \bar{I}$ into a pencil of the form shown in (3.6), where the sequence of indices satisfies

$$
\epsilon_{p} \geq \ldots \geq \epsilon_{2} \geq \epsilon_{1} \geq 1
$$

and $\left[B_{2}-\mu J_{2}\right]$ is a pencil for which the equation

$$
\left[B_{2}-\mu J_{2}\right] \hat{y}(\mu)=0
$$

has no nonzero solution, i.e., the columns are linearly independent.
By the same argument as in the proof of Lemma 1, Assumption 2 implies that there is also no vector $\hat{\varphi}(\mu) \neq 0$ such that

$$
\hat{\varphi}(\mu)^{\prime}\left[B_{2}-\mu J_{2}\right]=0,
$$

i.e., the rows of the pencil are also linearly independent. Hence $\left[B_{2}-\mu J_{2}\right]$ must be a square pencil of some dimension $q \times q$. (Note that it is possible that $q=0$, i.e., that $B_{2}$ and $J_{2}$ are null matrices.)

Adding up the columns of the matrix in (3.6), one observes that

$$
\sum_{i=1}^{p} \epsilon_{i}+q=n
$$

Adding up the rows, one similarly observes that

$$
\sum_{i=1}^{p}\left(\epsilon_{i}+1\right)+q=m
$$

from which it follows that the number of $M_{i}$ blocks must equal $p=m-n$.
This theorem implies that there exist nonsingular square matrices $P, Q$ of dimensions $n \times n$ and $m \times m$ respectively that satisfy (3.6).

## A. 6 Proof of Lemma 4

Suppose that (3.2) holds for any period $t>t_{0}$, and suppose that Assumption 2 holds. Lemma 3 then implies that there exist matrices $P, Q$ that define the decomposition (3.6), i.e., a decomposition of the form (2.20). Assumption 4 then allows us to decompose conditions (3.2) into separate subsystems as well. It follows from (2.21) that

$$
\begin{aligned}
{\left[\begin{array}{l}
\hat{\varphi}_{1, t} \\
\hat{\varphi}_{2, t}
\end{array}\right] } & =P\left[\begin{array}{cc}
T_{1}^{\prime} & 0 \\
0 & T_{2}^{\prime}
\end{array}\right]\left[\begin{array}{cc}
\left(T_{1}^{-1}\right)^{\prime} & 0 \\
0 & \left(T_{2}^{-1}\right)^{\prime}
\end{array}\right]\left[\begin{array}{l}
\tilde{\varphi}_{1, t} \\
\tilde{\varphi}_{2, t}
\end{array}\right] \\
& =\left[\begin{array}{cccc}
U_{1} & 0 & U_{2} & 0 \\
0 & X_{1} & 0 & X_{2}
\end{array}\right]\left[\begin{array}{c}
\left(T_{1}^{-1}\right)^{\prime} \tilde{\varphi}_{1, t} \\
\left(T_{2}^{-1}\right)^{\prime} \\
\tilde{\varphi}_{2, t}
\end{array}\right]
\end{aligned}
$$

so that

$$
\begin{align*}
\hat{\varphi}_{1, t} & =\left[\begin{array}{ll}
U_{1} & 0
\end{array}\right]\left(T_{1}^{-1}\right)^{\prime} \tilde{\varphi}_{1, t}+\left[\begin{array}{ll}
U_{2} & 0
\end{array}\right]\left(T_{2}^{-1}\right)^{\prime} \tilde{\varphi}_{2, t}  \tag{A.45}\\
\hat{\varphi}_{2, t} & =\left[\begin{array}{ll}
0 & X_{1}
\end{array}\right]\left(T_{1}^{-1}\right)^{\prime} \tilde{\varphi}_{1, t}+\left[\begin{array}{ll}
0 & X_{2}
\end{array}\right]\left(T_{2}^{-1}\right)^{\prime} \tilde{\varphi}_{2, t} . \tag{A.46}
\end{align*}
$$

Equations (A.45) and (A.46) respectively imply that

$$
\begin{align*}
& \hat{\varphi}_{1, t}-E_{t-1} \hat{\varphi}_{1, t}=\left[\begin{array}{cc}
U_{1} & 0
\end{array}\right]\left(T_{1}^{-1}\right)^{\prime}\left(\tilde{\varphi}_{1, t}-E_{t-1} \tilde{\varphi}_{1, t}\right)  \tag{А.47}\\
& \hat{\varphi}_{2, t}-E_{t-1} \hat{\varphi}_{2, t}=\left[\begin{array}{ll}
0 & X_{1}
\end{array}\right]\left(T_{1}^{-1}\right)^{\prime}\left(\tilde{\varphi}_{1, t}-E_{t-1} \tilde{\varphi}_{1, t}\right) . \tag{A.48}
\end{align*}
$$

Recall from Assumption 4 that both [ $\left.\begin{array}{ll}U_{1} & U_{2}\end{array}\right]$ and [ $\left.\begin{array}{ll}X_{1} & X_{2}\end{array}\right]$ are invertible matrices, and define the matrices $V_{1}, V_{2}, W_{1}$, and $W_{2}$ as in (3.10). Note that it follows from these definitions that

$$
\begin{array}{rllll}
V_{1} U_{1} & =I_{k_{1}}, & V_{1} U_{2}=0, & V_{2} U_{1}=0, & V_{2} U_{2}=I_{\tilde{n}-k_{1}} \\
W_{1} X_{1} & =I_{k_{2}}, & W_{1} X_{2}=0, & W_{2} X_{1}=0, & W_{2} X_{2}=I_{q-k_{2}} \tag{A.50}
\end{array}
$$

Premultiplying (A.47) by $V_{2}$ yields (3.11); premultiplying (A.48) by $W_{2}$ yields (3.12). Hence (3.11)-(3.12) must hold for all $t>t_{0}$.

Conversely, suppose that (3.11)-(3.12) hold in some period $t>t_{0}$. Using (A.45), (3.11) implies that

$$
\left[\begin{array}{ll}
I & 0
\end{array}\right]\left(T_{2}^{-1}\right)^{\prime}\left(\tilde{\varphi}_{2, t}-E_{t-1} \tilde{\varphi}_{2, t}\right)=0
$$

Similarly, using (A.46), (3.12) implies that

$$
\left[\begin{array}{ll}
0 & I
\end{array}\right]\left(T_{2}^{-1}\right)^{\prime}\left(\tilde{\varphi}_{2, t}-E_{t-1} \tilde{\varphi}_{2, t}\right)=0 .
$$

Together, these conditions imply that

$$
\left(T_{2}^{-1}\right)^{\prime}\left(\tilde{\varphi}_{2, t}-E_{t-1} \tilde{\varphi}_{2, t}\right)=0
$$

which implies (3.2). Hence (3.2) must hold for all $t>t_{0}$.

## A. 7 Proof of Proposition 3

In order to prove Proposition 3, we make use of a further preliminary result.
Lemma 12 Suppose that there exist matrices $P, Q$ that define a decomposition of the form (2.20), and that Assumptions 2, 4 and 5 hold. Then the $q \times q$ matrix

$$
\begin{equation*}
\left[\right] \tag{A.51}
\end{equation*}
$$

is invertible.

Proof. Let $x$ be an arbitrary vector of length $q$, partitioned as in (A.40). By Assumption 5 , knowing the values of the vectors $X_{1}^{\prime} J_{2} R_{2}^{\prime} x$ and $[0 I] N_{2}^{\prime} J_{2} R_{2}^{\prime} x$ allows one to reconstruct the entire vector $J_{2} R_{2}^{\prime} x$, and hence all elements of

$$
N_{2}^{\prime} J_{2} R_{2}^{\prime} x=\left[\begin{array}{c}
x_{1} \\
H^{\prime} x_{2}
\end{array}\right] .
$$

Since the elements of [0I] $N_{2}^{\prime} J_{2} R_{2}^{\prime} x=H^{\prime} x_{2}$ provide information only about the elements of $x_{2}$, it must be that each of the $k_{2}$ independent directions of variation of the elements of $x_{1}$ affects the elements of $X_{1}^{\prime} J_{2} R_{2}^{\prime} x$ in an independent direction. Thus Assumption 5 implies that the $k_{2} \times k_{2}$ matrix

$$
\Xi \equiv X_{1}^{\prime} J_{2} R_{2}^{\prime}\left[\begin{array}{c}
I_{k_{2}}  \tag{A.52}\\
0
\end{array}\right]
$$

is invertible.
Next, let $\hat{\varphi}$ be another arbitrary vector of length $q$, and let

$$
\check{\varphi} \equiv N_{2}^{-1} \hat{\varphi} \equiv\left[\begin{array}{c}
\check{\varphi}_{1} \\
\check{\varphi}_{2}
\end{array}\right] .
$$

One observes that

$$
R_{2} J_{2}^{\prime} \hat{\varphi}=\left[R_{2} J_{2}^{\prime} N_{2}\right] \check{\varphi}=\left[\begin{array}{cc}
I & 0 \\
0 & H
\end{array}\right] \check{\varphi}=\left[\begin{array}{c}
\check{\varphi}_{1} \\
H \check{\varphi}_{2}
\end{array}\right]
$$

Then in the case that

$$
\begin{equation*}
\hat{\varphi}=X_{1} f \tag{A.53}
\end{equation*}
$$

for some vector $f$ of length $k_{2}$, it follows that $\check{\varphi}_{1}=\Xi^{\prime} f$, where $\Xi$ is defined in (A.52).
Because $\Xi$ (and hence $\Xi^{\prime}$ ) is invertible, if for any $\hat{\varphi}$ of the form (A.53), $\check{\varphi}_{1}=0$, it follows that $f=0$ and hence that $\hat{\varphi}=0$. But it follows from (A.50) that a vector $\hat{\varphi}$ has a representation of the form (A.53) if and only if $W_{2} \hat{\varphi}=0$. Hence if any vector $\hat{\varphi}$ satisfies both $W_{2} \hat{\varphi}=0$ and $\check{\varphi}_{1}=0$, it must satisfy $\hat{\varphi}=0$. Alternatively, if it satisfies both the linear restrictions $W_{2} \hat{\varphi}=0$ and $\left[\begin{array}{ll}I & 0\end{array}\right] R_{2} J_{2}^{\prime} \hat{\varphi}=0$, it must be a zero vector. It follows from this that the matrix (A.51) must be invertible.

We may now proceed with the proof of Proposition 3. Let $\left\{\hat{\tau}_{2, t}\right\}$ be any process satisfying $\left\|\hat{\tau}_{2}\right\|<\infty$. Because (3.6) is a decomposition of the form (2.20), Lemma 1 guarantees that if Assumption 2 and 4 are satisfied, then there exist non-singular matrices $N_{2}, R_{2}$ such that (2.29) holds. Then premultiplying (3.9) by $R_{2}$ and using (2.29) yields

$$
\begin{align*}
\check{\varphi}_{1 t} & =\beta G E_{t} \check{\varphi}_{1, t+1}-\beta \check{\tau}_{1 t}  \tag{A.54}\\
E_{t} \check{\varphi}_{2, t+1} & =\left(\beta^{-1} H\right) \check{\varphi}_{2, t}+\check{\tau}_{2 t} \tag{A.55}
\end{align*}
$$

where

$$
\check{\varphi}_{t} \equiv\left[\begin{array}{c}
\check{\varphi}_{1 t} \\
\check{\varphi}_{2 t}
\end{array}\right] \equiv N_{2}^{-1} \hat{\varphi}_{2 t}, \quad\left[\begin{array}{c}
\check{\tau}_{1 t} \\
\check{\tau}_{2 t}
\end{array}\right] \equiv R_{2} \hat{\tau}_{2 t} .
$$

Because all eigenvalues of $\beta G$ have modulus less than 1, (A.54) has a solution for $\left\{\check{\varphi}_{1 t}\right\}$ such that $\left\|\check{\varphi}_{1}\right\|<\infty$, given by

$$
\begin{equation*}
\check{\varphi}_{1 t}=-\beta \sum_{j=0}^{\infty}(\beta G)^{j} E_{t} \check{\tau}_{1, t+j} . \tag{A.56}
\end{equation*}
$$

Then in any period $t$, the value of

$$
\check{\varphi}_{1 t}=\left[\begin{array}{ll}
I & 0
\end{array}\right] R_{2} J_{2}^{\prime} \hat{\varphi}_{2 t}
$$

is given by (A.56), while the value of $W_{2} \hat{\varphi}_{2 t}$ is given by (3.12). Thus the values of both $\check{\varphi}_{1 t}$ and $W_{2} \hat{\varphi}_{2 t}$ are given as functions of variables that are exogenous and/or predetermined in period $t$. But Lemma 12 implies that the mapping from the linear space of vectors $\hat{\varphi}$ to the values of $\check{\varphi}_{1}$ and $W_{2} \hat{\varphi}$ is an isomorphism, so this system of equations can be uniquely solved for the value of $\hat{\varphi}_{2 t}$. We thus obtain a unique solution for $\hat{\varphi}_{2 t}$ as a linear function of $E_{t-1} \hat{\varphi}_{2 t}$ and the $E_{t} \check{\tau}_{1, t+j}$ for $j \geq 0$.

This solution for $\hat{\varphi}_{2 t}$ allows us to solve for $\check{\varphi}_{2 t}$, and substituting this into (A.55) yields a value for $E_{t} \check{\varphi}_{2, t+1}$ as a linear function of $E_{t-1} \hat{\varphi}_{2 t}, \check{\tau}_{2 t}$, and the $E_{t} \check{\tau}_{1, t+j}$ for $j \geq 0$. The solution (A.56) implies that

$$
E_{t} \check{\varphi}_{1, t+1}=-\beta \sum_{j=0}^{\infty}(\beta G)^{j} E_{t} \check{\tau}_{1, t+j+1}
$$

Hence we can solve for the complete vector $E_{t} \check{\varphi}_{t+1}$ as a linear function of $E_{t-1} \hat{\varphi}_{2 t}$ and the exogenous state. Alternatively, we can solve for $E_{t} \hat{\varphi}_{2, t+1}=N_{2} E_{t} \check{\varphi}_{t+1}$ as a linear function of $E_{t-1} \hat{\varphi}_{2 t}$ and the exogenous state. Thus starting from an initial condition $E_{t_{0}-1} \hat{\varphi}_{2 t_{0}}$, we can solve for $\hat{\varphi}_{2 t_{0}}$ and $E_{t_{0}} \hat{\varphi}_{2, t_{0}+1}$; using this solution we can solve for $\hat{\varphi}_{2, t_{0}+1}$ and $E_{t_{0}+1} \hat{\varphi}_{2, t_{0}+2}$; and so on
recursively.
Thus it is possible to solve for the complete evolution $\left\{\hat{\varphi}_{2 t}\right\}$ for all $t \geq t_{0}$, given the initial condition $E_{t_{0}-1} \hat{\varphi}_{2 t_{0}}$ and the evolution of the exogenous state. By construction (A.54) and (A.55) are satisfied for each $t \geq t_{0}$, which implies that (3.9) is satisfied for each $t \geq t_{0}$. Likewise, (3.12) is satisfied for each $t>t_{0}$ by construction. Thus we obtain a process $\left\{\hat{\varphi}_{2 t}\right\}$ that satisfies both (3.9) for all $t \geq t_{0}$ and (3.12) for all $t>t_{0}$. Moreover, because all eigenvalues of $\beta^{-1} H$ have modulus less than 1, (A.55) implies that the constructed solution satisfies $\left\|\hat{\varphi}_{2}\right\|<\infty$.

## A. 8 Proof of Lemma 5

Consider the case of unidimensional policy so that $i=p=1$, and thus $\epsilon_{i}=\tilde{n}$.
Suppose that (3.8) hold for all $t \geq t_{0}$. The first element of the vector of FOCs (3.8) can be written

$$
\begin{equation*}
\beta^{-1} \hat{\varphi}_{1, t}^{(1)}=\hat{\tau}_{1, t}^{(1)} \tag{A.57}
\end{equation*}
$$

while the $j$-th element (for $2 \leq j \leq \tilde{n}$ ) can be written as

$$
\begin{equation*}
\beta^{-1} \hat{\varphi}_{1, t}^{(j)}=\hat{\tau}_{1, t}^{(j)}-E_{t} \hat{\varphi}_{1, t+1}^{(j-1)} . \tag{A.58}
\end{equation*}
$$

This system of equations can be solved recursively for the $\left\{\hat{\varphi}_{1, t}^{(j)}\right\}$, yielding

$$
\begin{equation*}
\hat{\varphi}_{1, t}^{(j)}=\beta E_{t}\left[\Gamma^{j}\left(\beta L^{-1}\right) \hat{\tau}_{1, t}\right] \tag{A.59}
\end{equation*}
$$

for each $1 \leq j \leq \tilde{n}$, where $\Gamma^{j}(\mu)$ is the $j$-th row of the matrix polynomial $\Gamma(\mu)$. This gives the vector of conditions (3.16).

The $(\tilde{n}+1)$-st element of the vector of FOCs (3.8) states that

$$
\begin{equation*}
E_{t} \hat{\varphi}_{1, t+1}^{(\tilde{n})}=\hat{\tau}_{1, t}^{(\tilde{n}+1)} \tag{A.60}
\end{equation*}
$$

Substituting solution (A.59) for $\hat{\varphi}_{1, t}^{(j)}$, we obtain

$$
\begin{equation*}
\hat{\tau}_{1, t}^{(\tilde{n}+1)}=E_{t}\left[\beta L^{-1} \Gamma^{\tilde{n}}\left(\beta L^{-1}\right) \hat{\tau}_{1, t}\right] \tag{A.61}
\end{equation*}
$$

which implies (3.17). Here we use the fact that

$$
\delta(\mu)^{\prime}=(-\mu)^{\tilde{n}}\left[e_{\tilde{n}+1}^{\prime}-\mu^{-1} \Gamma^{\tilde{n}}\left(\mu^{-1}\right)\right]
$$

where $e_{\tilde{n}+1}^{\prime}$ is a $1 \times(\tilde{n}+1)$ vector of the form $e_{\tilde{n}+1}^{\prime} \equiv\left[\begin{array}{lll}0 \ldots 0 & 1\end{array}\right]$. Thus both (3.16) and (3.17) must hold in all periods $t \geq t_{0}$.

To prove the converse, suppose that the processes $\left\{\hat{\tau}_{1, t}\right\}$ and $\left\{\hat{\varphi}_{1, t}\right\}$ satisfy (3.16) and (3.17) in all periods $t \geq t_{0}$. Condition (3.16) implies (A.59) for each $1 \leq j \leq \tilde{n}$, which in turn implies condition (A.57), and condition (A.58) for each $2 \leq j \leq \tilde{n}$. These are just the first $\tilde{n}$ elements of the vector of FOCs (3.8). Condition (3.17) implies (A.61), which together with the case $j=\tilde{n}$ of (A.59) implies (A.60). This is just the $(\tilde{n}+1)$-st element of the vector of FOCs (3.8). Thus the entire vector of conditions (3.8) must hold in each period $t \geq t_{0}$.

## A. 9 Proof of Proposition 4

To prove Proposition 4, it will be useful to appeal to the following Lemma.
Lemma 13 Suppose that the processes $\left\{\hat{\tau}_{1, t}\right\}$ and $\left\{\hat{\varphi}_{1, t}\right\}$ satisfy (3.16) for all $t \geq t_{0}$. Then for any $1 \leq j \leq \tilde{n}$, and any $t \geq t_{0}$,

$$
\begin{equation*}
E_{t} z_{t+j-1}-E_{t-1} z_{t+j-1}=-(-\beta)^{-j}\left[\hat{\varphi}_{1, t}^{(j)}-E_{t-1} \hat{\varphi}_{1, t}^{(j)}\right] \tag{А.62}
\end{equation*}
$$

where $\hat{\varphi}_{1, t}^{(j)}$ is the $j$-th element of the vector $\hat{\varphi}_{1, t}$. When $t=t_{0}$, the expression $E_{t_{0}-1} \hat{\varphi}_{1, t_{0}}^{(j)}$ is taken to refer to the historical expectations $\beta E_{t_{0}-1}\left[\Gamma^{j}\left(\beta L^{-1}\right) \hat{\tau}_{1, t_{0}}\right]$.

Proof. Using the definition (3.15) we have

$$
\begin{aligned}
E_{t} z_{t+j-1}-E_{t-1} z_{t+j-1}= & E_{t}\left[\hat{\tau}_{1, t+j-1}^{(1)}-\beta^{-1} \hat{\tau}_{1, t+j-2}^{(2)}+\ldots+\left(-\beta^{-1}\right)^{j-1} \hat{\tau}_{1, t}^{(j)}\right] \\
& -E_{t-1}\left[\hat{\tau}_{1, t+j-1}^{(1)}-\beta^{-1} \hat{\tau}_{1, t+j-2}^{(2)}+\ldots+\left(-\beta^{-1}\right)^{j-1} \hat{\tau}_{1, t}^{(j)}\right] \\
= & \left(-\beta^{-1}\right)^{j-1}\left\{E_{t}\left[\Gamma^{j}\left(\beta L^{-1}\right) \hat{\tau}_{1, t}\right]-E_{t-1}\left[\Gamma^{j}\left(\beta L^{-1}\right) \hat{\tau}_{1, t}\right]\right\} \\
= & -(-\beta)^{-j}\left[\hat{\varphi}_{1, t}^{(j)}-E_{t-1} \hat{\varphi}_{1, t}^{(j)}\right] .
\end{aligned}
$$

Here, the first equality uses the definition of $\delta(\mu)^{\prime}$, and the fact that $E_{t} \hat{\tau}_{1, t-k}=E_{t-1} \hat{\tau}_{1, t-k}$ for any $k \geq 1$. The second equality uses the definition of $\Gamma(\mu)$, denotes by $\Gamma^{j}(\mu)$ its $j$ th row. The third equality uses (3.16). In the case that $t=t_{0}$, the replacement of $\beta E_{t-1}\left[\Gamma^{j}\left(\beta L^{-1}\right) \hat{\tau}_{1, t}\right]$ by $E_{t-1} \hat{\varphi}_{1, t}^{(j)}$ is justified under the definition of $E_{t_{0}-1} \hat{\varphi}_{1, t_{0}}^{(j)}$ proposed above.

We may now proceed with the proof of Proposition 4. Condition (3.8) implies that (3.16) and (3.17) must hold for all $t \geq t_{0}$, using Lemma 5. The fact that (3.16) holds implies that (A.62) must also hold for all $t \geq t_{0}$, using Lemma 13.

Let us first consider any period $t \geq t_{0}+\tilde{n}-k_{1}$. Then

$$
\begin{align*}
E_{t} z_{t+k_{1}} & =\sum_{j=1}^{\tilde{n}-k_{1}}\left(E_{t+1-j} z_{t+k_{1}}-E_{t-j} z_{t+k_{1}}\right)+E_{t-\left(\tilde{n}-k_{1}\right)} z_{t+k_{1}} \\
& =\sum_{j=1}^{\tilde{n}-k_{1}}(-\beta)^{-\left(k_{1}+j\right)}\left[\hat{\varphi}_{1, t+1-j}^{\left(k_{1}+j\right)}-E_{t-j} \hat{\varphi}_{1, t+1-j}^{\left(k_{1}+j\right)}\right] . \tag{A.63}
\end{align*}
$$

Here the second line uses (A.62) to replace the first term on the right-hand side of the first line, and uses (3.17) to eliminate the second term.

Given Assumption 6, (3.11) implies that the entire vector $\left[\hat{\varphi}_{1, t}-E_{t-1} \hat{\varphi}_{1, t}\right]$ can be reconstructed from its first $k_{1}$ elements, using

$$
\hat{\varphi}_{1, t}-E_{t-1} \hat{\varphi}_{1, t}=\left[\begin{array}{c}
I_{k_{1}}  \tag{A.64}\\
\Phi
\end{array}\right]\left[\bar{\varphi}_{1, t}-E_{t-1} \bar{\varphi}_{1, t}\right]
$$

for any $t \geq t_{0}+1$, where $\bar{\varphi}_{1, t}$ is the vector consisting of the first $k_{1}$ elements of $\hat{\varphi}_{1, t}$. Using (A.64) to substitute for the terms on the right-hand side of (A.63), we obtain

$$
\begin{align*}
E_{t} z_{t+k_{1}} & =\sum_{j=1}^{\tilde{n}-k_{1}}-(-\beta)^{-\left(k_{1}+j\right)} \phi_{j}^{\prime}\left[\bar{\varphi}_{1, t+1-j}-E_{t-j} \bar{\varphi}_{1, t+1-j}\right] \\
& =\sum_{j=1}^{\tilde{n}-k_{1}}(-\beta)^{-\left(k_{1}+j\right)} \phi_{j}^{\prime} w_{t+1-j} \\
& =(-\beta)^{-k_{1}} \sum_{j=1}^{\tilde{n}-k_{1}} \phi_{j}^{\prime} \Omega_{t}^{j}  \tag{A.65}\\
& =(-\beta)^{-k_{1}} \operatorname{tr}\left[\Phi \Omega_{t}\right]
\end{align*}
$$

which establishes (3.20). Here we use the notation $\phi_{j}^{\prime}$ for the $j$-th row of $\Phi$ and the notation $\Omega_{t}^{j}$ for the $j$-th column of $\Omega_{t}$. In addition, the second line uses Lemma 13 to substitute for the elements of $\bar{\varphi}_{1, t+1-j}-E_{t-j} \bar{\varphi}_{1, t+1-j}$, and definition (3.18), while the third line uses the definition of $\Omega_{t}$.

Let us now consider any period $t_{0} \leq t<t_{0}+\tilde{n}-k_{1}$. Then

$$
\begin{align*}
E_{t} z_{t+k_{1}} & =\sum_{j=1}^{t-t_{0}}\left(E_{t+1-j} z_{t+k_{1}}-E_{t-j} z_{t+k_{1}}\right)+E_{t_{0}} z_{t+k_{1}} \\
& =(-\beta)^{-k_{1}} \sum_{j=1}^{t-t_{0}} \phi_{j}^{\prime} \Omega_{t}^{j}+E_{t_{0}} z_{t+k_{1}} \tag{A.66}
\end{align*}
$$

where the second line is obtained using the same reasoning as was used to derive (A.63) and (A.65).

For the given historical expectations $e_{t_{0}-1}$, let $\Xi_{1, t_{0}-1}$ be given by (3.21) where $\chi_{t_{0}-1}$ is the vector whose $j$-th element is given by (3.22). With this definition, (3.13) together with (3.8) and Lemma 5 imply that

$$
V_{2}\left[\hat{\varphi}_{1, t_{0}}\right]=V_{2} E_{t_{0}-1}\left[\hat{\varphi}_{1, t_{0}}\right]+V_{22} \chi_{t_{0}-1}
$$

Premultiplying by $V_{22}^{-1}$ and noting that $\Phi \equiv-V_{22}^{-1} V_{21}$, this yields

$$
\left[-\Phi \quad I_{\tilde{n}-k_{1}}\right]\left[\hat{\varphi}_{1, t_{0}}-E_{t_{0}-1} \hat{\varphi}_{1, t_{0}}\right]=\chi_{t_{0}-1}
$$

which can alternatively be written

$$
\begin{equation*}
\left[\Phi \quad\left(-I_{\tilde{n}-k_{1}}\right)\right] w_{t_{0}}=\chi_{t_{0}-1} \tag{А.67}
\end{equation*}
$$

using Lemma 13 and definition (3.18).
For any $1 \leq j \leq \tilde{n}-k_{1}$, the $j$-th row of (A.67) can be written

$$
\begin{aligned}
& \phi_{j}^{\prime} \bar{w}_{t_{0}}-(-\beta)^{k_{1}+j}\left[E_{t_{0}} z_{t_{0}+k_{1}+j-1}-E_{t_{0}-1} z_{t_{0}+k_{1}+j-1}\right] \\
= & (-\beta)^{k_{1}+j} E_{t_{0}-1} z_{t_{0}+k_{1}+j-1}-\sum_{i=1}^{\tilde{n}-k_{1}-j}(-\beta)^{-i} \phi_{j+i}^{\prime} \bar{w}_{t_{0}-i},
\end{aligned}
$$

or alternatively,

$$
(-\beta)^{k_{1}+j} E_{t_{0}} z_{t_{0}+k_{1}+j-1}=\sum_{i=0}^{\tilde{n}-k_{1}-j}(-\beta)^{-i} \phi_{j+i}^{\prime} \bar{w}_{t_{0}-i}=\sum_{h=j}^{\tilde{n}-k_{1}}(-\beta)^{j} \phi_{h}^{\prime} \Omega_{t_{0}+j-1}^{h}
$$

Thus if we let $j=t+1-t_{0}$, we find that

$$
E_{t_{0}} z_{t+k_{1}}=(-\beta)^{-k_{1}} \sum_{h=t-t_{0}+1}^{\tilde{n}-k_{1}} \phi_{h}^{\prime} \Omega_{t}^{h}
$$

Using this to substitute for the final term on the right-hand side of (A.66), we obtain

$$
E_{t} z_{t+k_{1}}=(-\beta)^{-k_{1}} \sum_{j=1}^{\tilde{n}-k_{1}} \phi_{j}^{\prime} \Omega_{t}^{j}
$$

so that (3.20) is satisfied for each $t_{0} \leq t<t_{0}+\tilde{n}-k_{1}$. Since we have already shown that (3.20) holds for any $t \geq t_{0}+\tilde{n}-k_{1}$, it follows that (3.20) is satisfied for each $t \geq t_{0}$.

## A. 10 Proof of Lemma 6

For period $t_{0}$, conditions (3.23) and (3.24) hold by assumption, given the initial condition (3.13) and the initial Lagrange multipliers (3.21). Next, (3.8) implies $\hat{\varphi}_{1, t}=\beta E_{t}\left[\Gamma\left(\beta L^{-1}\right) \hat{\tau}_{1, t}\right]$ at all dates $t \geq t_{0}$, by Lemma 5 . Using this, conditions (3.11) in turn imply that

$$
V_{2} \hat{\varphi}_{1, t}=V_{2} E_{t-1} \hat{\varphi}_{1, t}=V_{2} \beta E_{t-1}\left[\Gamma\left(\beta L^{-1}\right) \hat{\tau}_{1, t}\right]
$$

for all $t>t_{0}$. Equation (3.23) thus holds in all period $t \geq t_{0}$, where $\Xi_{1, t-1}$ is given by (3.24) in period $t=t_{0}$, and

$$
\Xi_{1, t-1}=V_{2} \beta E_{t-1}\left[\Gamma\left(\beta L^{-1}\right) \hat{\tau}_{1, t}\right]
$$

for all $t>t_{0}$. To prove that (3.24) holds at all dates, we need to show that $V_{22} \chi_{t-1}=0$ for all periods $t>t_{0}$.

Given the initial Lagrange multipliers $\Xi_{1, t_{0}}$ defined in (3.21) and using Proposition 4 implies that (3.20) must hold at all dates $t \geq t_{0}$. Condition (3.20) then implies that for any $t \geq t_{0}$ and any $1 \leq j \leq \tilde{n}-k_{1}$,

$$
\begin{align*}
& E_{t} z_{t+k_{1}+j-1}=E_{t}\left[E_{t+j-1} z_{t+k_{1}+j-1}\right]=(-\beta)^{-k_{1}} \operatorname{tr}\left[\Phi E_{t} \Omega_{t+j-1}\right] \\
&=(-\beta)^{-k_{1}} E_{t}\left[(-\beta)^{-1} \phi_{1}^{\prime} \bar{w}_{t+j-1}+\ldots+(-\beta)^{-\left(\tilde{n}-k_{1}\right)} \phi_{\tilde{n}-k_{1}}^{\prime} \bar{w}_{t+j+k_{1}-\tilde{n}}\right] \\
&=(-\beta)^{-k_{1}}\left[(-\beta)^{-j} \phi_{j}^{\prime} \bar{w}_{t}+\ldots+(-\beta)^{-\left(\tilde{n}-k_{1}\right)} \phi_{\tilde{n}-k_{1}}^{\prime} \bar{w}_{t+j+k_{1}-\tilde{n}}\right] \\
&=(-\beta)^{-\left(k_{1}+j\right)} \sum_{i=0}^{n}-k_{1}-j  \tag{A.68}\\
&
\end{align*}
$$

Using this to substitute for $E_{t-1} z_{t+k_{1}+j-1}$ in (3.22), we obtain

$$
\chi_{t-1}^{j}=\phi_{j}^{\prime} E_{t-1} \bar{w}_{t}=0
$$

for all $t>t_{0}$, and any $1 \leq j \leq \tilde{n}-k_{1}$. It follows that $\chi_{t-1}=0$, and hence that $V_{22} \chi_{t-1}=0$ for all $t>t_{0}$.

## A. 11 Proof of Proposition 5

Let the process $\left\{\hat{\varphi}_{1, t}\right\}$ be given by (3.16) for all $t \geq t_{0}$. Note that if $k_{1}<\tilde{n}$, satisfaction of (3.20) for all $t \geq t_{0}$ implies that

$$
E_{t} z_{t+\tilde{n}}=E_{t}\left[E_{t+\tilde{n}-k_{1}} z_{t+\tilde{n}}\right]=(-\beta)^{-k_{1}} \operatorname{tr}\left[\Phi E_{t} \Omega_{t+\tilde{n}-k_{1}}\right]=0
$$

for any $t \geq t_{0}$, using the fact that $E_{t} \bar{w}_{t+j}=0$ for any $t \geq t_{0}$ and any $j \geq 1$. But if $k_{1}=\tilde{n}$, $\Omega_{t}$ is a null matrix, and (3.20) also implies that $E_{t} z_{t+\tilde{n}}=0$ in this case as well. Hence the target criterion (3.20) implies condition (3.17). Then by Lemma 5, the fact that (3.16) and (3.17) hold for all $t \geq t_{0}$ implies that (3.8) holds for all $t \geq t_{0}$.

Recall from (A.68) that condition (3.20) also implies

$$
E_{t} z_{t+k_{1}+j-1}=(-\beta)^{-\left(k_{1}+j\right)} \sum_{i=0}^{\tilde{n}-k_{1}-j}(-\beta)^{-i} \phi_{j+i}^{\prime} \bar{w}_{t-i}
$$

for any $t \geq t_{0}$ and any $1 \leq j \leq \tilde{n}-k_{1}$. It follows that for all $t>t_{0}$ and all $1 \leq j \leq \tilde{n}-k_{1}$ :

$$
E_{t} z_{t+k_{1}+j-1}-E_{t-1} z_{t+k_{1}+j-1}=(-\beta)^{-\left(k_{1}+j\right)} \phi_{j}^{\prime} \bar{w}_{t}
$$

so that

$$
w_{t} \equiv\left[\begin{array}{c}
(-\beta)\left[z_{t}-E_{t-1} z_{t}\right] \\
\vdots \\
(-\beta)^{k_{1}}\left[E_{t} z_{t+k_{1}-1}-E_{t-1} z_{t+k_{1}-1}\right] \\
(-\beta)^{k_{1}+1}\left[E_{t} z_{t+k_{1}}-E_{t-1} z_{t+k_{1}}\right] \\
\vdots \\
(-\beta)^{\tilde{n}}\left[E_{t} z_{t+\tilde{n}-1}-E_{t-1} z_{t+\tilde{n}-1}\right]
\end{array}\right]=\left[\begin{array}{c}
I_{k_{1}} \\
\Phi
\end{array}\right] \bar{w}_{t} .
$$

Premultiplying by $V_{2}$ implies in turn that

$$
V_{2} w_{t}=\left[\begin{array}{ll}
V_{21} & V_{22}
\end{array}\right]\left[\begin{array}{c}
I_{k_{1}} \\
\Phi
\end{array}\right] \bar{w}_{t}=0
$$

for any $t>t_{0}$. For the process $\left\{\hat{\varphi}_{1, t}\right\}$ given by (3.16) for all $t \geq t_{0}$, we can apply Lemma 13 to express $w_{t}$ as

$$
w_{t}=-\left[\hat{\varphi}_{1, t}-E_{t-1} \hat{\varphi}_{1, t}\right]
$$

for all $t \geq t_{0}$. Since $V_{2} w_{t}=0$ for any $t>t_{0}$, equation (3.11) must hold for all $t>t_{0}$.
To show that the initial condition (3.13) also holds, where the vector $\Xi_{1, t_{0}-1}$ is defined by (3.21), note that (A.68) implies

$$
(-\beta)^{k_{1}+j} E_{t_{0}} z_{t_{0}+k_{1}+j-1}=\sum_{i=0}^{\tilde{n}-k_{1}-j}(-\beta)^{-i} \phi_{j+i}^{\prime} \bar{w}_{t_{0}-i}
$$

for any $1 \leq j \leq \tilde{n}-k_{1}$. Subtracting $\phi_{j}^{\prime} \bar{w}_{t_{0}}+(-\beta)^{k_{1}+j} E_{t_{0}-1} z_{t_{0}+k_{1}+j-1}$ on both sides (where again expectations taken at date $t_{0}-1$ denote historical forecasts) yields

$$
-\phi_{j}^{\prime} \bar{w}_{t_{0}}+(-\beta)^{k_{1}+j}\left[E_{t_{0}} z_{t_{0}+k_{1}+j-1}-E_{t_{0}-1} z_{t_{0}+k_{1}+j-1}\right]=-\chi_{t_{0}-1}^{j}
$$

where $\chi_{t_{0}-1}^{j}$ is the $j$-th element of the vector $\chi_{t_{0}-1}$, defined in (3.22). Since the previous expression holds for any $1 \leq j \leq \tilde{n}-k_{1}$, we may rewrite it in matrix form as

$$
\left[\begin{array}{cc}
-\Phi & I_{\tilde{n}-k_{1}}
\end{array}\right] w_{t_{0}}=-\chi_{t_{0}-1}
$$

using definition (3.18). Using again definition (3.18) and Lemma 13 we can equivalently rewrite this as

$$
\left[-\Phi \quad I_{\tilde{n}-k_{1}}\right]\left[\hat{\varphi}_{1, t_{0}}-E_{t_{0}-1} \hat{\varphi}_{1, t_{0}}\right]=\chi_{t_{0}-1}
$$

or as

$$
V_{2} \hat{\varphi}_{1, t_{0}}=V_{2} E_{t_{0}-1}\left[\hat{\varphi}_{1, t_{0}}\right]+V_{22} \chi_{t_{0}-1},
$$

after premultiplying on both sides by $V_{22}$. Using (3.16) to replace $\hat{\varphi}_{1, t_{0}}$ with $\beta E_{t_{0}}\left[\Gamma\left(\beta L^{-1}\right) \hat{\tau}_{1, t_{0}}\right]$ on the right-hand side yields the initial condition (3.13), where the vector $\Xi_{1, t_{0}-1}$ is defined by (3.21).

## B Linearization of the FOCs

In this section, we provide a linearization of the first-order conditions (1.8).

## B. 1 Linearizing $\left(D_{1} \pi\left(y_{t}, \xi_{t}\right)\right)^{\prime}$.

We note that $\left(D_{1} \pi\left(y_{t}, \xi_{t}\right)\right)^{\prime}$ is a $(m \times 1)$ vector-valued function. A first-order approximation of this vector function yields

$$
\left(D_{1} \pi\left(y_{t}, \xi_{t}\right)\right)^{\prime}=\left(D_{1} \pi\right)^{\prime}+D_{1}\left[\left(D_{1} \pi\right)^{\prime}\right]\left(y_{t}-\bar{y}\right)+D_{2}\left[\left(D_{1} \pi\right)^{\prime}\right]\left(\xi_{t}-\bar{\xi}\right)+\mathcal{O}\left(\epsilon^{2}\right)
$$

## B. 2 Linearizing $\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t}$.

To approximate the vector functions $\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t}$, recall that the matrix function $\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime}$ is $(m \times k)$ and $\theta_{t}$ is a $(k \times 1)$ vector. It will be convenient to rewrite $\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t}$ as follows

$$
\begin{aligned}
\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t} & =\operatorname{vec}\left[\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t}\right]=\operatorname{vec}\left[I_{m}\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t}\right] \\
& =\left(\theta_{t}^{\prime} \otimes I_{m}\right) \operatorname{vec}\left[\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime}\right] .
\end{aligned}
$$

We can then write the $(m \times k m)$ matrix $\left(\theta_{t}^{\prime} \otimes I_{m}\right)$ as

$$
\begin{aligned}
\left(\theta_{t}^{\prime} \otimes I_{m}\right) & =\left(\bar{\theta}^{\prime} \otimes I_{m}\right)+\left(\theta_{t}^{\prime} \otimes I_{m}\right)-\left(\bar{\theta}^{\prime} \otimes I_{m}\right) \\
& =\left(\bar{\theta}^{\prime} \otimes I_{m}\right)+\left(\tilde{\theta}_{t}^{\prime} \otimes I_{m}\right)
\end{aligned}
$$

where $\tilde{\theta}_{t}=\theta_{t}-\bar{\theta}$. Next, the $(k m \times 1)$ function $\operatorname{vec}\left[\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime}\right]$ can be approximated as follows

$$
\begin{aligned}
\operatorname{vec}\left[\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime}\right]= & \operatorname{vec}\left[\left(D_{1} F\right)^{\prime}\right]+\frac{\partial v e c\left[\left(D_{1} F\right)^{\prime}\right]}{\partial y_{t}^{\prime}}\left(y_{t}-\bar{y}\right)+\frac{\partial v e c\left[\left(D_{1} F\right)^{\prime}\right]}{\partial \xi_{t}^{\prime}}\left(\xi_{t}-\bar{\xi}\right) \\
& +\frac{\partial v e c\left[\left(D_{1} F\right)^{\prime}\right]}{\partial y_{t}^{\prime}}\left(y_{t-1}-\bar{y}\right)+\mathcal{O}\left(\epsilon^{2}\right)
\end{aligned}
$$

which can be written as
$\operatorname{vec}\left[\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime}\right]=\operatorname{vec}\left[\left(D_{1} F\right)^{\prime}\right]+D_{1}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t}+D_{3}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t-1}+D_{2}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{\xi}_{t}+\mathcal{O}\left(\epsilon^{2}\right)$.
A first-order approximation of $\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t}$ is thus given by

$$
\begin{aligned}
& \left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t} \\
& =\left[\left(\bar{\theta}^{\prime} \otimes I_{m}\right)+\left(\tilde{\theta}_{t}^{\prime} \otimes I_{m}\right)\right] \times\left\{\operatorname{vec}\left[\left(D_{1} F\right)^{\prime}\right]+D_{1}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t}\right. \\
& \left.+D_{3}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t-1}+D_{2}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{\xi}_{t}+\mathcal{O}\left(\epsilon^{2}\right)\right\} \\
& =\left(\bar{\theta}^{\prime} \otimes I_{m}\right)\left\{\operatorname{vec}\left[\left(D_{1} F\right)^{\prime}\right]+D_{1}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t}+D_{3}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t-1}+D_{2}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{\xi}_{t}\right\} \\
& +\left(\tilde{\theta}_{t}^{\prime} \otimes I_{m}\right)\left\{\operatorname{vec}\left[\left(D_{1} F\right)^{\prime}\right]+D_{1}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t}+D_{3}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t-1}+D_{2}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{\xi}_{t}\right\} \\
& +\mathcal{O}\left(\epsilon^{2}\right)
\end{aligned}
$$

Note that several elements of the last two lines are of order $\mathcal{O}\left(\epsilon^{2}\right)$. They can thus be omitted as part of the approximation error. Furthermore, as $\left(\bar{\theta}^{\prime} \otimes I_{m}\right)$ vec $\left[\left(D_{1} F\right)^{\prime}\right]=v e c\left[I_{m}\left(D_{1} F\right)^{\prime} \bar{\theta}\right]$ $=\left(D_{1} F\right)^{\prime} \bar{\theta}$, we can rewrite

$$
\begin{aligned}
\left(D_{1} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)^{\prime} \theta_{t}= & \left(\bar{\theta}^{\prime} \otimes I_{m}\right)\left\{D_{1}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t}+D_{3}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t-1}+D_{2}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{\xi}_{t}\right\} \\
& +\left(D_{1} F\right)^{\prime} \bar{\theta}+\left(D_{1} F\right)^{\prime} \tilde{\theta}_{t}+\mathcal{O}\left(\epsilon^{2}\right)
\end{aligned}
$$

## B. 3 Linearization of the FOCs

Having linearized the two vectors described above, we are now prepared to provide a firstorder approximation to (1.7):

$$
\begin{aligned}
0= & D_{1}\left[\left(D_{1} \pi\right)^{\prime}\right] \tilde{y}_{t}+D_{2}\left[\left(D_{1} \pi\right)^{\prime}\right] \tilde{\xi}_{t} \\
& +\left(D_{1} F\right)^{\prime} \tilde{\theta}_{t}+\left(\bar{\theta}^{\prime} \otimes I_{m}\right)\left\{D_{1}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t}+D_{3}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{y}_{t-1}+D_{2}\left[\left(D_{1} F\right)^{\prime}\right] \tilde{\xi}_{t}\right\} \\
& +\beta E_{t}\left\{\left(D_{3} F\right)^{\prime} \tilde{\theta}_{t+1}+\left(\bar{\theta}^{\prime} \otimes I_{m}\right)\left\{D_{1}\left[\left(D_{3} F\right)^{\prime}\right] \tilde{y}_{t+1}+D_{3}\left[\left(D_{3} F\right)^{\prime}\right] \tilde{y}_{t}+D_{2}\left[\left(D_{3} F\right)^{\prime}\right] \tilde{\xi}_{t+1}\right\}\right\} \\
& +\left(D_{1} g\right)^{\prime} \tilde{\Theta}_{t}+\left(\bar{\Theta}^{\prime} \otimes I_{m}\right)\left\{D_{1}\left[\left(D_{1} g\right)^{\prime}\right] \tilde{y}_{t}+D_{3}\left[\left(D_{1} g\right)^{\prime}\right] E_{t} \tilde{y}_{t+1}+D_{2}\left[\left(D_{1} g\right)^{\prime}\right] \tilde{\xi}_{t}\right\} \\
& +\beta^{-1}\left(D_{3} g\right)^{\prime} \tilde{\Theta}_{t-1}+\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) \beta^{-1}\left\{D_{1}\left[\left(D_{3} g\right)^{\prime}\right] \tilde{y}_{t-1}+D_{3}\left[\left(D_{3} g\right)^{\prime}\right] \tilde{y}_{t}+D_{2}\left[\left(D_{3} g\right)^{\prime}\right] \tilde{\xi}_{t-1}\right\} \\
& +\left(D_{1} \pi\right)^{\prime}+\left(D_{1} F\right)^{\prime} \bar{\theta}+\beta\left(D_{3} F\right)^{\prime} \bar{\theta}+\left(D_{1} g\right)^{\prime} \bar{\Theta}+\beta^{-1}\left(D_{3} g\right)^{\prime} \bar{\Theta}+\mathcal{O}\left(\epsilon^{2}\right) .
\end{aligned}
$$

Using the fact that in steady state

$$
0=\left(D_{1} \pi\right)^{\prime}+\left(D_{1} F\right)^{\prime} \bar{\theta}+\beta\left(D_{3} F\right)^{\prime} \bar{\theta}+\left(D_{1} g\right)^{\prime} \bar{\Theta}+\beta^{-1}\left(D_{3} g\right)^{\prime} \bar{\Theta}
$$

the last line reduces to the approximation error.
Let us define

$$
\begin{aligned}
S \equiv & D_{1}\left[\left(D_{1} \pi\right)^{\prime}\right]+\left(\bar{\theta}^{\prime} \otimes I_{m}\right)\left\{D_{1}\left[\left(D_{1} F\right)^{\prime}\right]+\beta D_{3}\left[\left(D_{3} F\right)^{\prime}\right]\right\} \\
& +\left(\bar{\Theta}^{\prime} \otimes I_{m}\right)\left\{D_{1}\left[\left(D_{1} g\right)^{\prime}\right]+\beta^{-1} D_{3}\left[\left(D_{3} g\right)^{\prime}\right]\right\} \\
R \equiv & \left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{3}\left[\left(D_{1} F\right)^{\prime}\right]+\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) \beta^{-1} D_{1}\left[\left(D_{3} g\right)^{\prime}\right] \\
R^{\dagger} \equiv & \left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{3} F\right)^{\prime}\right]+\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) \beta^{-1} D_{3}\left[\left(D_{1} g\right)^{\prime}\right]
\end{aligned}
$$

and the matrix polynomial

$$
\begin{aligned}
B(L) \equiv & \left\{D_{2}\left[\left(D_{1} \pi\right)^{\prime}\right]+\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{1} F\right)^{\prime}\right]+\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{1} g\right)^{\prime}\right]\right\} \cdot L \\
& +\beta^{-1}\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{3} g\right)^{\prime}\right] \cdot L^{2}+\beta\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{3} F\right)^{\prime}\right]
\end{aligned}
$$

Using this, we can rewrite the approximation to the first-order conditions as:

$$
\begin{aligned}
0= & S \tilde{y}_{t}+R \tilde{y}_{t-1}+\beta R^{\dagger} E_{t} \tilde{y}_{t+1}+E_{t}\left[B(L) \tilde{\xi}_{t+1}\right] \\
& +\left(D_{3} F\right)^{\prime} \beta E_{t} \tilde{\theta}_{t+1}+\left(D_{1} g\right)^{\prime} \tilde{\Theta}_{t}+\left(D_{1} F\right)^{\prime} \tilde{\theta}_{t}+\beta^{-1}\left(D_{3} g\right)^{\prime} \tilde{\Theta}_{t-1}+\mathcal{O}\left(\epsilon^{2}\right) \\
= & S \tilde{y}_{t}+R \tilde{y}_{t-1}+\beta R^{\dagger} E_{t} \tilde{y}_{t+1}+E_{t}\left[B(L) \tilde{\xi}_{t+1}\right]+E_{t}\left[\left(\bar{A}-\beta^{-1} \bar{I} L\right)^{\prime} \tilde{\varphi}_{t+1}\right]+\mathcal{O}\left(\epsilon^{2}\right) .
\end{aligned}
$$

These conditions further reduce to (2.16), using the following lemma.
Lemma 14 If $\pi(), F()$ and $g()$ are twice continuously differentiable, then $S$ is a symmetric matrix (i.e., $S=S^{\prime}$ ) and $R^{\dagger}=R^{\prime}$.

Proof. We start by showing that

$$
\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{3} F\right)^{\prime}\right]=\left\{\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{3}\left[\left(D_{1} F\right)^{\prime}\right]\right\}^{\prime}
$$

Expanding the derivatives, we have

$$
D_{1}\left[\left(D_{3} F\right)^{\prime}\right]=\frac{\partial v e c\left[\left(\frac{\partial F}{\partial y_{t-1}^{\prime}}\right)^{\prime}\right]}{\partial y_{t}^{\prime}}=\frac{\partial v e c\left[\begin{array}{ccc}
\partial F_{1} / \partial y_{1, t-1} & \cdots & \partial F_{k} / \partial y_{1, t-1} \\
\vdots & & \vdots \\
\partial F_{1} / \partial y_{m, t-1} & \cdots & \partial F_{k} / \partial y_{m, t-1}
\end{array}\right]}{\partial y_{t}^{\prime}}
$$

so that

$$
D_{1}\left[\left(D_{3} F\right)^{\prime}\right]=\left[\begin{array}{ccc}
\frac{\partial^{2} F_{1}}{\partial y_{1, t-1} \partial y_{1, t}} & \cdots & \frac{\partial^{2} F_{1}}{\partial y_{1, t-1} \partial y_{m, t}} \\
\vdots & & \vdots \\
\frac{\partial^{2} F_{1}}{\partial y_{m, t-1} \partial y_{1, t}} & \cdots & \frac{\partial^{2} F_{1}}{\partial y_{m, t-} \partial y_{m, t}} \\
\vdots & & \vdots \\
\frac{\partial^{2} F_{k}}{\partial y_{1, t-1} \partial y_{1, t}} & \cdots & \frac{\partial^{2} F_{k}}{\partial y_{1, t-1} \partial y_{m, t}} \\
\vdots & & \vdots \\
\frac{\partial^{2} F_{k}}{\partial y_{m, t-1} \partial y_{1, t}} & \cdots & \frac{\partial^{2} F_{k}}{\partial y_{m, t-1} \partial y_{m, t}}
\end{array}\right] .
$$

By Young's theorem, if $F()$ is twice continuously differentiable, then $\partial^{2} F_{j} /\left(\partial y_{k, t-1} \partial y_{l, t}\right)=$ $\partial^{2} F_{j} /\left(\partial y_{l, t} \partial y_{k, t-1}\right)$ for all $j, k, l$. It follows that

$$
\begin{aligned}
& \left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{3} F\right)^{\prime}\right] \\
& =\left[\begin{array}{cccccccc}
\bar{\theta}_{1} & & 0 & & \bar{\theta}_{m} & & 0 \\
& \ddots & & \cdots & & \ddots & \\
0 & & \bar{\theta}_{1} & & 0 & & \bar{\theta}_{m}
\end{array}\right]\left[\begin{array}{cccc}
\frac{\partial^{2} F_{1}}{\partial y_{1, t} \partial y_{1, t-1}} & \cdots & \frac{\partial^{2} F_{1}}{\partial y_{m, t} \partial y_{1, t-1}} \\
\vdots & & \vdots \\
\frac{\partial^{2} F_{1}}{\partial y_{1, t} \partial y_{m, t-1}} & \cdots & \frac{\partial^{2} F_{1}}{\partial y_{m, t} \partial y_{m, t-1}} \\
\vdots & & & \vdots \\
\frac{\partial^{2} F_{k}}{\partial y_{1, t} \partial y_{1, t-1}} & \cdots & \frac{\partial^{2} F_{k}}{\partial y_{m, t} \partial y_{1, t-1}} \\
\vdots & & & \\
& & & \\
& & & \\
\frac{\partial^{2} F_{k}}{\partial y_{1, t} \partial y_{m, t-1}} & \cdots & & \\
& & & \\
& & & \\
\partial y_{m, t} \partial y_{k} \\
\partial_{m, t-1}
\end{array}\right]
\end{aligned}
$$

On the other hand:

$$
D_{3}\left[\left(D_{1} F\right)^{\prime}\right]=\frac{\partial v e c\left[\begin{array}{ccc}
\partial F_{1} / \partial y_{1, t} & \cdots & \partial F_{k} / \partial y_{1, t} \\
\vdots & & \vdots \\
\partial F_{1} / \partial y_{m, t} & \cdots & \partial F_{k} / \partial y_{m, t}
\end{array}\right]}{\partial y_{t-1}^{\prime}}
$$

so that using again Young's theorem, we have

$$
\begin{aligned}
& \left\{\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{3}\left[\left(D_{1} F\right)^{\prime}\right]\right\}^{\prime}
\end{aligned}
$$

$$
\begin{aligned}
& =\left[\begin{array}{ccc}
\bar{\theta}_{1} \frac{\partial^{2} F_{1}}{\partial y_{1, t} \partial y_{1, t-1}}+\cdots+\bar{\theta}_{m} \frac{\partial^{2} F_{k}}{\partial y_{1, t} \partial y_{1, t-1}} & \cdots & \bar{\theta}_{1} \frac{\partial^{2} F_{1}}{\partial y_{m, t} \partial y_{1, t-1}}+\cdots+\bar{\theta}_{m} \frac{\partial^{2} F_{k}}{\partial y_{m, t} \partial y_{1, t-1}} \\
\vdots & \bar{\theta}_{m} \frac{\partial^{2} F_{k}}{\partial y_{1, t} \partial y_{m, t-1}} & \cdots \\
\bar{\theta}_{1} \frac{\partial^{2} F_{1}}{\partial y_{1, t} \partial y_{m, t-1}}+\cdots+\frac{\partial^{2} F_{1}}{\partial y_{m, t} \partial y_{m, t-1}}+\cdots+\bar{\theta}_{m} \frac{\partial^{2} F_{k}}{\partial y_{m, t} \partial y_{m, t-1}}
\end{array}\right]
\end{aligned}
$$

which is equal to the expression derived before for $\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{3} F\right)^{\prime}\right]$.
Having shown that $\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{3} F\right)^{\prime}\right]=\left\{\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{3}\left[\left(D_{1} F\right)^{\prime}\right]\right\}^{\prime}$, it is then easy to see (when replacing $\bar{\theta}$ with $\bar{\Theta}$ and $F$ with $g$ ) that

$$
\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{3} g\right)^{\prime}\right]=\left\{\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) D_{3}\left[\left(D_{1} g\right)^{\prime}\right]\right\}^{\prime}
$$

It follows that $R^{\dagger}=R^{\prime}$.
A similar derivation can be performed to show that

$$
\begin{aligned}
\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{j}\left[\left(D_{j} F\right)^{\prime}\right] & =\left\{\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{j}\left[\left(D_{j} F\right)^{\prime}\right]\right\}^{\prime} \\
\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) D_{j}\left[\left(D_{j} g\right)^{\prime}\right] & =\left\{\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) D_{j}\left[\left(D_{j} g\right)^{\prime}\right]\right\}^{\prime}
\end{aligned}
$$

for $j=1,3$, and $D_{1}\left[\left(D_{1} \pi\right)^{\prime}\right]=\left\{D_{1}\left[\left(D_{1} \pi\right)^{\prime}\right]\right\}^{\prime}$. This implies that the matrix $S$ is symmetric.

## C Second-Order Conditions

We now perform a second-order approximation of the Lagrangian (1.6) around the optimal steady state. This involves a second-order approximation of each of the three terms in the square brackets of (1.6), evaluating the derivatives at the optimal steady state.

## C. 1 Second-Order Approximation of the Lagrangian

A second-order approximation of the function $\pi\left(y_{t}, \xi_{t}\right)$ yields

$$
\begin{aligned}
\pi\left(y_{t}, \xi_{t}\right)= & \pi(\bar{y}, \bar{\xi})+\left[\begin{array}{ll}
D_{1} \pi & D_{2} \pi
\end{array}\right]\left[\begin{array}{c}
\tilde{y}_{t} \\
\tilde{\xi}_{t}
\end{array}\right] \\
& +\frac{1}{2}\left[\begin{array}{ll}
\tilde{y}_{t}^{\prime} & \tilde{\xi}_{t}^{\prime}
\end{array}\right]\left[\begin{array}{ll}
D_{1}\left[\left(D_{1} \pi\right)^{\prime}\right] & D_{2}\left[\left(D_{1} \pi\right)^{\prime}\right] \\
D_{1}\left[\left(D_{2} \pi\right)^{\prime}\right] & D_{2}\left[\left(D_{2} \pi\right)^{\prime}\right]
\end{array}\right]\left[\begin{array}{l}
\tilde{y}_{t} \\
\tilde{\xi}_{t}
\end{array}\right]+\mathcal{O}\left(\epsilon^{3}\right) .
\end{aligned}
$$

A second-order approximation of the function $\theta_{t}^{\prime} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)$ yields

$$
\begin{aligned}
\theta_{t}^{\prime} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)= & \bar{\theta}^{\prime}\left[\begin{array}{lll}
D_{1} F & D_{2} F & D_{3} F
\end{array}\right]\left[\begin{array}{c}
\tilde{y}_{t} \\
\tilde{\xi}_{t} \\
\tilde{y}_{t-1}
\end{array}\right] \\
& +\frac{1}{2}\left[\begin{array}{llll}
\tilde{y}_{t}^{\prime} & \tilde{\xi}_{t}^{\prime} & \tilde{y}_{t-1}^{\prime} & \tilde{\theta}_{t}^{\prime}
\end{array}\right] \mathcal{H}\left(\begin{array}{c}
\left.\bar{\theta}^{\prime} F\right)\left[\begin{array}{c}
\tilde{y}_{t} \\
\tilde{\xi}_{t} \\
\tilde{y}_{t-1} \\
\tilde{\theta}_{t}
\end{array}\right]+\mathcal{O}\left(\epsilon^{3}\right)
\end{array}\right.
\end{aligned}
$$

where the Hessian matrix

$$
\mathcal{H}\left(\bar{\theta}^{\prime} F\right)=\left[\begin{array}{cccc}
\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{1} F\right)^{\prime}\right] & \left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{1} F\right)^{\prime}\right] & \left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{3}\left[\left(D_{1} F\right)^{\prime}\right] & \left(D_{1} F\right)^{\prime} \\
\left(\bar{\theta}^{\prime} \otimes I_{n_{\xi}}\right) D_{1}\left[\left(D_{2} F\right)^{\prime}\right] & \left(\bar{\theta}^{\prime} \otimes I_{n_{\xi}}\right) D_{2}\left[\left(D_{2} F\right)^{\prime}\right] & \left(\bar{\theta}^{\prime} \otimes I_{n_{\xi}}\right) D_{3}\left[\left(D_{2} F\right)^{\prime}\right] & \left(D_{2} F\right)^{\prime} \\
\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{3} F\right)^{\prime}\right] & \left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{3} F\right)^{\prime}\right] & \left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{3}\left[\left(D_{3} F\right)^{\prime}\right] & \left(D_{3} F\right)^{\prime} \\
D_{1} F & D_{2} F & D_{3} F & 0
\end{array}\right]
$$

has $j i$ blocks of the form

$$
\mathcal{H}_{j i}\left(\theta_{t}^{\prime} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)=D_{j}\left\{\left[D_{i}\left(\theta_{t}^{\prime} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)\right]^{\prime}\right\}
$$

evaluated at the optimal steady state, where the indices $j, i=1, \ldots 4$, refer respectively to the vectors $y_{t}, \xi_{t}, y_{t-1}, \theta_{t}$, and where $n_{\xi}$ is the dimension of the vector $\xi_{t}$. This approximation involves neither a constant term nor a first-order term in $\tilde{\theta}_{t}$, as we use the fact that $F(\bar{y}, \bar{\xi} ; \bar{y})=0$ in the optimal steady state. To obtain the derivatives $\mathcal{H}_{j i}\left(\theta_{t}^{\prime} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right)$, for $j, i=1,2,3$, we use

$$
\begin{aligned}
\mathcal{H}_{j i}\left(\theta_{t}^{\prime} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right) & =D_{j}\left\{\left[\theta_{t}^{\prime} D_{i} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right]^{\prime}\right\}=D_{j}\left\{\left[D_{i} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right]^{\prime} \theta_{t}\right\} \\
& =D_{j}\left\{\operatorname{vec}\left(\left[D_{i} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right]^{\prime} \theta_{t}\right)\right\}=D_{j}\left\{\left(\theta_{t}^{\prime} \otimes I\right) \cdot \operatorname{vec}\left(\left[D_{i} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right]^{\prime}\right)\right\} \\
& =\left(\theta_{t}^{\prime} \otimes I\right) \cdot D_{j}\left\{\left[D_{i} F\left(y_{t}, \xi_{t} ; y_{t-1}\right)\right]^{\prime}\right\} .
\end{aligned}
$$

To obtain second-order approximation of $\Theta_{t-1}^{\prime} g\left(y_{t-1}, \xi_{t-1} ; y_{t}\right)$, we proceed similarly, replac$\operatorname{ing} F()$ with $g()$ and $\theta_{t}$ with $\Theta_{t-1}$.

The second-order approximation of the Lagrangian is thus given by

$$
\begin{aligned}
\mathcal{L}_{t_{0}}= & E_{t_{0}}\left\{\sum _ { t = t _ { 0 } } ^ { \infty } \beta ^ { t - t _ { 0 } } \left(\pi(\bar{y}, \bar{\xi})+\left[D_{1} \pi, D_{2} \pi\right]\left[\begin{array}{c}
\tilde{y}_{t} \\
\tilde{\xi}_{t}
\end{array}\right]+\frac{1}{2}\left[\tilde{y}_{t}^{\prime}, \tilde{\xi}_{t}^{\prime}\right]\left[\begin{array}{c}
D_{1}\left[\left(D_{1} \pi\right)^{\prime}\right] \\
D_{1}\left[\left(D_{2} \pi\right)^{\prime}\right]
\end{array} \begin{array}{c}
D_{2}\left[\left(D_{1} \pi\right)^{\prime}\right. \\
D_{2}\left[\left(D_{2} \pi\right)^{\prime}\right]
\end{array}\right]\left[\begin{array}{c}
\tilde{y}_{t} \\
\tilde{\xi}_{t}
\end{array}\right]\right.\right. \\
& +\bar{\theta}^{\prime}\left[D_{1} F, D_{2} F, D_{3} F\right]\left[\begin{array}{c}
\tilde{y}_{t} \\
\tilde{\xi}_{t} \\
\tilde{y}_{t-1}
\end{array}\right]+\frac{1}{2}\left[\tilde{y}_{t}^{\prime}, \tilde{\xi}_{t}^{\prime}, \tilde{y}_{t-1}^{\prime}, \tilde{\theta}_{t}^{\prime}\right] \mathcal{H}\left(\bar{\theta}^{\prime} F\right)\left[\begin{array}{c}
\tilde{y}_{t} \\
\tilde{\xi}_{t} \\
\tilde{y}_{t-1} \\
\tilde{\theta}_{t}
\end{array}\right] \\
& \left.\left.+\beta^{-1} \bar{\Theta}^{\prime}\left[D_{1} g, D_{2} g, D_{3} g\right]\left[\begin{array}{c}
\tilde{y}_{t-1} \\
\tilde{\xi}_{t-1} \\
\tilde{y}_{t}
\end{array}\right]+\frac{\beta^{-1}}{2}\left[\tilde{y}_{t-1}^{\prime}, \tilde{\xi}_{t-1}^{\prime}, \tilde{y}_{t}^{\prime}, \tilde{\Theta}_{t-1}^{\prime}\right] \mathcal{H}\left(\bar{\Theta}^{\prime} g\right)\left[\begin{array}{c}
\tilde{y}_{t-1} \\
\tilde{\xi}_{t-1} \\
\tilde{y}_{t} \\
\tilde{\Theta}_{t-1}
\end{array}\right]\right)\right\} \\
& +\mathcal{O}\left(\epsilon^{3}\right)
\end{aligned}
$$

Expanding this expression, recognizing that each quadratic term is a scalar so that it is equal to its transpose, and using the definitions of the matrices $\bar{A}, \bar{I}$ and $\bar{C}$ in (2.8), and of $S, R$ and the matrix polynomial $B(L)$ in (2.16), we can rewrite the Lagrangian as

$$
\begin{aligned}
\mathcal{L}_{t_{0}}= & E_{t_{0}}\left\{\sum _ { t = t _ { 0 } } ^ { \infty } \beta ^ { t - t _ { 0 } } \left(\pi(\bar{y}, \bar{\xi})+\left[D_{2} \pi+\bar{\theta}^{\prime}\left(D_{2} F\right)\right] \tilde{\xi}_{t}+\beta^{-1} \bar{\Theta}^{\prime}\left(D_{2} g\right) \tilde{\xi}_{t-1}\right.\right. \\
& +\frac{1}{2} \tilde{\xi}_{t}^{\prime}\left(D_{2}\left[\left(D_{2} \pi\right)^{\prime}\right]+\left(\bar{\theta}^{\prime} \otimes I_{n_{\xi}}\right) D_{2}\left[\left(D_{2} F\right)^{\prime}\right]\right) \tilde{\xi}_{t}+\frac{\beta^{-1}}{2} \tilde{\xi}_{t-1}^{\prime}\left(\bar{\Theta}^{\prime} \otimes I_{n_{\xi}}\right) D_{2}\left[\left(D_{2} g\right)^{\prime}\right] \tilde{\xi}_{t-1} \\
& +\left[D_{1} \pi+\bar{\varphi}^{\prime}\left(\bar{A}-\beta^{-1} \bar{I}\right)\right] \tilde{y}_{t} \\
& \left.\left.+\frac{1}{2}\left[\tilde{y}_{t}^{\prime} S \tilde{y}_{t}+\tilde{y}_{t}^{\prime} R \tilde{y}_{t-1}+\tilde{y}_{t-1}^{\prime} R^{\prime} \tilde{y}_{t}\right]+\tilde{y}_{t}^{\prime}\left[B(L) \tilde{\xi}_{t+1}\right]+\left[\tilde{\varphi}_{t+1}^{\prime} \bar{A}-\beta^{-1} \tilde{\varphi}_{t}^{\prime} \bar{I}\right] \tilde{y}_{t}+\tilde{\varphi}_{t+1}^{\prime} \bar{C} \tilde{\xi}_{t+1}\right)\right\} \\
& +\beta^{-1} \tilde{\varphi}_{t_{0}}^{\prime} \bar{A} \tilde{y}_{t_{0}-1}+\tilde{\theta}_{t_{0}}^{\prime}\left(D_{2} F\right) \tilde{\xi}_{t_{0}}+\beta^{-1} \tilde{\Theta}_{t_{0}-1}^{\prime}\left(D_{2} g\right) \tilde{\xi}_{t_{0}-1} \\
& +\beta^{-1} \bar{\varphi}^{\prime} \bar{A} \tilde{y}_{t_{0}-1}+\frac{1}{2} \tilde{y}_{t_{0}-1}^{\prime}\left\{\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{3}\left[\left(D_{3} F\right)^{\prime}\right]+\beta^{-1}\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) D_{1}\left[\left(D_{1} g\right)^{\prime}\right]\right\} \tilde{y}_{t_{0}-1} \\
& +\tilde{y}_{t_{0}-1}^{\prime}\left\{\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{3} F\right)^{\prime}\right] \tilde{\xi}_{t_{0}}+\beta^{-1}\left(\bar{\Theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{1} g\right)^{\prime}\right] \tilde{\xi}_{t_{0}-1}\right\}+\mathcal{O}\left(\epsilon^{3}\right) .
\end{aligned}
$$

In deriving this expression, we use the properties of the transposes discussed in the proof of Lemma 14, so that, e.g.,

$$
\left(\bar{\theta}^{\prime} \otimes I_{n_{\xi}}\right) D_{1}\left[\left(D_{2} F\right)^{\prime}\right]=\left\{\left(\bar{\theta}^{\prime} \otimes I_{m}\right) D_{2}\left[\left(D_{1} F\right)^{\prime}\right]\right\}^{\prime}
$$

In this approximation of the Lagrangian, we note that the terms in the first two lines are either a constant, or functions only of exogenous disturbances, so that they are independent of the path of endogenous variables (hence of the policy chosen at date $t_{0}$ ). Furthermore, the steady-state condition (1.12) implies that the term in square brackets in the third line is equal to zero, thereby cancelling the terms that are linear in $\tilde{y}_{t}$. In addition, the last two rows involve terms independent of any policy chosen at date $t_{0}$, and terms of third order or
smaller. It follows that the Lagrangian can be rewritten as

$$
\begin{aligned}
\mathcal{L}_{t_{0}}= & E_{t_{0}}\left\{\sum_{t=t_{0}}^{\infty} \beta^{t-t_{0}}\left[\tilde{y}_{t}^{\prime}\left(\frac{1}{2} S \tilde{y}_{t}+R \tilde{y}_{t-1}+\left[B(L) \tilde{\xi}_{t+1}\right]+\left(\bar{A}-\beta^{-1} \bar{I} L\right)^{\prime} \tilde{\varphi}_{t+1}\right)+\tilde{\varphi}_{t+1}^{\prime} \bar{C} \tilde{\xi}_{t+1}\right]\right\} \\
& +\beta^{-1} \tilde{\varphi}_{t_{0}}^{\prime} \bar{A}^{\prime} \tilde{y}_{t_{0}-1}+\tilde{\theta}_{t_{0}}^{\prime}\left(D_{2} F\right) \tilde{\xi}_{t_{0}}+\beta^{-1} \tilde{\Theta}_{t_{0}-1}^{\prime}\left(D_{2} g\right) \tilde{\xi}_{t_{0}-1}+\operatorname{tip}+\mathcal{O}\left(\epsilon^{3}\right)
\end{aligned}
$$

where tip denotes terms independent of policy chosen at date $t_{0}$.

## C. 2 Perturbation of the Lagrangian Around the Optimal Steady State

As discussed in section 2.2.1, we consider a perturbation of the path $\left\{y_{t}\right\}$ from its optimal steady state equilibrium, maintaining the other vectors $\xi_{t}, \theta_{t}, \Theta_{t}$ unchanged. The perturbation is such that $\tilde{y}_{t}^{\varepsilon}=\tilde{y}_{t}$ for all $t<\tau$, and an arbitrary date $\tau \geq t_{0}$, while $\tilde{y}_{\tau+i}^{\varepsilon}=\tilde{y}_{\tau+i}+\varepsilon y_{i}$ for all $i \geq 0$. Here $\left\{y_{i}\right\}$ is a bounded sequence of vectors of dimension $m$, not all equal to zero, such that

$$
\begin{equation*}
\bar{I} y_{0}=0, \quad \bar{A} y_{i}-\bar{I} y_{i+1}=0 \quad \text { for all } i \geq 0 \tag{C.1}
\end{equation*}
$$

and $\varepsilon$ is a random quantity, the value of which is determined only at date $\tau$. (In the case that $\varepsilon$ is determined at date $\tau=t_{0}$, then it turns out that $\varepsilon$ is a deterministic quantity).

The perturbed path $\left\{\tilde{y}_{t}^{\varepsilon}\right\}$ remains consistent with the linearized structural equations (2.8)-(2.9) as long as the random variable $\varepsilon$ has conditional mean zero as of date $\tau-1$ (regardless of the state of the world at that date). If we let $\varepsilon$ be some mean-zero bounded random variable with unit variance multiplied by a scale factor $|\varepsilon|$, then for any small enough value of $|\varepsilon|$, then this also approximates a feasible perturbation under the exact structural relations (1.2)-(1.4), up to an error of order $\mathcal{O}\left(|\varepsilon|^{2}\right)$.

In the perturbed equilibrium, the Lagrangian is equal to

$$
\begin{aligned}
\mathcal{L}_{t_{0}}^{\varepsilon}= & E_{t_{0}}\left\{\sum_{t=t_{0}}^{\tau-1} \beta^{t-t_{0}}\left[\tilde{y}_{t}^{\prime}\left(\frac{1}{2} S \tilde{y}_{t}+R \tilde{y}_{t-1}+\left[B(L) \tilde{\xi}_{t+1}\right]+\left(\bar{A}-\beta^{-1} \bar{I} L\right)^{\prime} \tilde{\varphi}_{t+1}\right)+\tilde{\varphi}_{t+1}^{\prime} \bar{C} \tilde{\xi}_{t+1}\right]\right\} \\
& +E_{t_{0}}\left\{\beta^{t-t_{0}}\left(\tilde{y}_{\tau}+\varepsilon y_{0}\right)^{\prime} R \tilde{y}_{\tau-1}\right\} \\
& +E_{t_{0}}\left\{\sum _ { i = 0 } ^ { \infty } \beta ^ { \tau + i - t _ { 0 } } \left[( \tilde { y } _ { \tau + i } + \varepsilon y _ { i } ) ^ { \prime } \left(\frac{1}{2} S\left(\tilde{y}_{\tau+i}+\varepsilon y_{i}\right)+\beta R^{\prime}\left(\tilde{y}_{\tau+i+1}+\varepsilon y_{i+1}\right)+\left[B(L) \tilde{\xi}_{\tau+i+1}\right]\right.\right.\right. \\
& \left.\left.\left.+\left(\bar{A}-\beta^{-1} \bar{I} L\right)^{\prime} \tilde{\varphi}_{\tau+i+1}\right)+\tilde{\varphi}_{\tau+i+1}^{\prime} \bar{C} \tilde{\xi}_{\tau+i+1}\right]\right\} \\
& +\beta^{-1} \tilde{\varphi}_{t_{0}}^{\prime} \bar{A}^{\prime} \tilde{y}_{t_{0}-1}+\tilde{\theta}_{t_{0}}^{\prime}\left(D_{2} F\right) \tilde{\xi}_{t_{0}}+\beta^{-1} \tilde{\Theta}_{t_{0}-1}^{\prime}\left(D_{2} g\right) \tilde{\xi}_{t_{0}-1}+t i p+\mathcal{O}\left(|\epsilon, \varepsilon|^{3}\right) .
\end{aligned}
$$

The increase in the Lagrangian due to the perturbation can thus be expressed as

$$
\begin{aligned}
\mathcal{L}_{t_{0}}^{\varepsilon}-\mathcal{L}_{t_{0}}= & E_{t_{0}}\left\{\sum _ { i = 0 } ^ { \infty } \beta ^ { \tau + i - t _ { 0 } } \left(\frac{1}{2}\left[y_{i}^{\prime} S y_{i}+y_{i+1}^{\prime} \beta R y_{i}+y_{i}^{\prime} \beta R^{\prime} y_{i+1}\right] \varepsilon^{2}\right.\right. \\
& \left.\left.+y_{i}^{\prime}\left[S \tilde{y}_{\tau+i}+\beta R^{\prime} \tilde{y}_{\tau+i+1}+R \tilde{y}_{\tau+i-1}+\left(\bar{A}-\beta^{-1} \bar{I} L\right)^{\prime} \tilde{\varphi}_{\tau+i+1}+B(L) \tilde{\xi}_{\tau+i+1}\right] \varepsilon\right)\right\} \\
& +\mathcal{O}\left(|\epsilon, \varepsilon|^{3}\right) .
\end{aligned}
$$

Since the linearized first-order conditions (2.16) imply that the term in square brackets in the second line is of second order, this term becomes of third order once multiplied by $\varepsilon$, so that the increase in the Lagrangian due to the perturbation reduces to

$$
\mathcal{L}_{t_{0}}^{\varepsilon}-\mathcal{L}_{t_{0}}=\frac{1}{2} \sum_{i=0}^{\infty} \beta^{\tau+i-t_{0}}\left[y_{i}^{\prime} S y_{i}+y_{i+1}^{\prime} \beta R y_{i}+y_{i}^{\prime} \beta R^{\prime} y_{i+1}\right]|\varepsilon|^{2}+\mathcal{O}\left(|\epsilon, \varepsilon|^{3}\right)
$$

## D Real Kronecker Canonical Form

The following theorem adapts Gantmacher's (1959) proof of Theorem 3 (Chap. 12, vol. 2) to the case of a real Kronecker canonical form of regular matrix pencils.

Theorem 7 Real Kronecker canonical form. Consider the matrix pencil $A-\mu \hat{I}$, with $A, \hat{I} \in$ $\mathbb{R}^{n \times n}$. Suppose that its characteristic polynomial $\operatorname{det}[A-\mu \hat{I}]$, of order $n-k$, for $0 \leq k \leq n$, is not identically zero. Then there exist non-singular matrices $\tilde{U}, \tilde{V} \in \mathbb{R}^{n \times n}$ satisfying

$$
\tilde{U}(A-\mu \hat{I}) \tilde{V}=\left[\begin{array}{cc}
I_{k} & 0 \\
0 & \tilde{G}
\end{array}\right]-\mu\left[\begin{array}{cc}
\tilde{H} & 0 \\
0 & I_{n-k}
\end{array}\right]
$$

where $\tilde{H} \in \mathbb{R}^{k \times k}$ is a nilpotent matrix of the Jordan form (i.e., with ones on the first super diagonal and zeros everywhere else), and $\tilde{G} \in \mathbb{R}^{(n-k) \times(n-k)}$ is a block-diagonal matrix of the real Jordan form. Each of the diagonal blocks of $\tilde{G}$ is either of the form

$$
\left[\begin{array}{ccccc}
\mu_{i} & 1 & 0 & & 0  \tag{D.1}\\
0 & \mu_{i} & 1 & \ddots & \\
& \ddots & \ddots & \ddots & 0 \\
& & \ddots & \ddots & 1 \\
0 & & & 0 & \mu_{i}
\end{array}\right]
$$

where $\mu_{i} \in \mathbb{R}$, or

$$
\left[\begin{array}{ccccc}
M_{i} & I_{2} & 0 & & 0  \tag{D.2}\\
0 & M_{i} & I_{2} & \ddots & \\
& \ddots & \ddots & \ddots & 0 \\
& & \ddots & \ddots & I_{2} \\
0 & & & 0 & M_{i}
\end{array}\right]
$$

where the submatrices $M_{i}$ are of the form

$$
M_{i}=\left[\begin{array}{cc}
\alpha_{i} & \beta_{i} \\
-\beta_{i} & \alpha_{i}
\end{array}\right]
$$

with $\alpha_{i}, \beta_{i} \in \mathbb{R}$. Furthermore, the $n-k$ finite eigenvalues (including their algebraic multiplicity) of the matrix pencil $A-\mu \hat{I}$ are equal to eigenvalues (including their algebraic multiplicity) of the matrix pencil $\tilde{G}-\mu I$.

Proof. Note that $P(\mu)=(A-\mu \hat{I})$ is a regular pencil since $A, \hat{I}$ are square matrices and $\operatorname{det}[A-\mu \hat{I}]$ is not identically zero. Then there exists a number $c \in \mathbb{R}$ such that $A_{1} \equiv A-c \hat{I}$ $\in \mathbb{R}^{n \times n}$ satisfies $\left|A_{1}\right| \neq 0$. We represent the given pencil in the form $P(\mu)=A-c \hat{I}-(\mu-c) \hat{I}$ $=A_{1}-(\mu-c) \hat{I}$, and multiply it on the left by $A_{1}^{-1}$ to obtain $A_{1}^{-1} P(\mu)=I-(\mu-c) A_{1}^{-1} \hat{I}$. By similarity transformation into a real Jordan canonical form (Laub (2005), p. 83), there exists an invertible matrix $T \in \mathbb{R}^{n \times n}$ such that $A_{1}^{-1} \hat{I}=T\left\{J_{0}, \tilde{J}\right\} T^{-1}$ where $\left\{J_{0}, \tilde{J}\right\} \in \mathbb{R}^{n \times n}$ denotes a block diagonal matrix in which $J_{0}$ is a nilpotent Jordan matrix, and $\tilde{J}=\left\{J_{1}, J_{2}, \ldots, J_{q}\right\}$ is a real, block diagonal matrix which satisfies $\operatorname{det} \tilde{J} \neq 0$, with each block being of the form (D.1) in the case that $\mu_{i}$ are real eigenvalues of $A_{1}^{-1} \hat{I}$, and of the form (D.2) in the case of complex conjugate eigenvalues $\mu_{i}=\alpha_{i} \pm \beta i$ of $A_{1}^{-1} \hat{I}$.

Using this we can put the pencil in the form

$$
A_{1}^{-1} P(\mu)=I-(\mu-c) T\left\{J_{0}, \tilde{J}\right\} T^{-1}
$$

or equivalently

$$
T^{-1} A_{1}^{-1} P(\mu) T=\left\{I+c J_{0}-\mu J_{0}, I+c \tilde{J}-\mu \tilde{J}\right\}
$$

Multiplying on the right by $\left\{\left(I+c J_{0}\right)^{-1}, \tilde{J}^{-1}\right\}$ we have ${ }^{25}$

$$
T^{-1} A_{1}^{-1} P(\mu) T\left\{\left(I+c J_{0}\right)^{-1}, \tilde{J}^{-1}\right\}=\left\{I-\mu\left(I+c J_{0}\right)^{-1} J_{0},\left(\tilde{J}^{-1}+c I\right)-\mu I\right\}
$$

By a similarity transformation, we can write $\left(I+c J_{0}\right)^{-1} J_{0}=T_{0} \tilde{H} T_{0}^{-1}$ where $T_{0}$ is a real matrix which satisfies $\left|T_{0}\right| \neq 0$ and where $\tilde{H}=\left\{H^{\left(u_{1}\right)}, H^{\left(u_{2}\right)}, \ldots H^{\left(u_{s}\right)}\right\}$ is a Jordan matrix in which the submatrices $H^{(u)}$ of order $u$ are have ones on the first superdiagonal and zeroes in all other elements. (This is because $J_{0}$, and hence $\left(I+c J_{0}\right)^{-1} J_{0}$, are nilpotent and that all eigenvalues of a nilpotent matrix are always 0 ).

Thus by multiplying the matrix pencil above by $\left\{T_{0}^{-1}, I\right\}$ on the left and $\left\{T_{0}, I\right\}$ on the right, we get

$$
\begin{aligned}
\tilde{U}(A-\mu \hat{I}) \tilde{V} & =\left\{T_{0}^{-1} T_{0}-\mu T_{0}^{-1} J_{0}\left(I+c J_{0}\right)^{-1} T_{0},\left(\tilde{J}^{-1}+c I\right)-\mu I\right\} \\
& =\{I-\mu \tilde{H}, \tilde{G}-\mu I\}
\end{aligned}
$$

where $\tilde{U} \equiv\left\{T_{0}^{-1}, I\right\} T^{-1} A_{1}^{-1}, \tilde{V} \equiv T\left\{\left(I+c J_{0}\right)^{-1}, \tilde{J}^{-1}\right\}\left\{T_{0}, I\right\}$, and $\tilde{G} \equiv \tilde{J}^{-1}+c I$ are real matrices. Given that the matrices $\tilde{U}$ and $\tilde{V}$ are non-singular, the matrix pencils $(A-\mu \hat{I})$ and $\{I-\mu \tilde{H}, \tilde{G}-\mu I\}$ are strictly equivalent (see Definition 1, Gantmacher, 1959, Chap. 12). It follows from Theorem 2 of Gantmacher (1959, Chap. 12) that these two matrix pencils have the same ("finite" and "infinite") elementary divisors. Since all $k$ infinite elementary divisors of $(A-\mu \hat{I})$ are associated with $I-\mu \tilde{H}$, and the $n-k$ finite elementary divisors are associated with $\tilde{G}-\mu I$, the $n-k$ finite eigenvalues (including their algebraic multiplicity) of the matrix pencil $A-\mu \hat{I}$ are equal to the eigenvalues (including their algebraic multiplicity) of the matrix pencil $\tilde{G}-\mu I$. Thus $\tilde{H}$ is of dimensions $k \times k$ while $\tilde{G}$ is $(n-k) \times(n-k)$.

[^3]We have thus shown that for any real regular pencil $A-\mu \hat{I}$, there exist real matrices $\tilde{U}, \tilde{V}$ such that

$$
\tilde{U} A \tilde{V}=\left[\begin{array}{cc}
I & 0 \\
0 & \tilde{G}
\end{array}\right], \quad \tilde{U} \hat{I} \tilde{V}=\left[\begin{array}{cc}
\tilde{H} & 0 \\
0 & I
\end{array}\right]
$$

where $\tilde{G}$ is an invertible matrix of the (real) Jordan form and $\tilde{H}$ is an nilpotent matrix of the Jordan form.

## E Target Criterion in Model of Section 4

## E. 1 Some Details on the Model

In this model, each household seeks to maximize its lifetime utility

$$
E_{t_{0}} \sum_{t=t_{0}}^{\infty} \beta^{t-t_{0}}\left[\tilde{u}\left(C_{t} ; \xi_{t}\right)-\int_{0}^{1} \tilde{v}\left(h_{t}(j) ; \xi_{t}\right) d j\right]
$$

where $C_{t} \equiv\left[\int_{0}^{1} c_{t}(i)^{(\theta-1) / \theta}\right]^{\theta /(\theta-1)}$ is a Dixit-Stiglitz aggregate of consumption of each of a continuum of differentiated goods, with an elasticity of substitution $\theta>1$, and $h_{t}(j)$ is the quantity supplied of labor of type $j$. Each differentiated good is supplied by a single monopolistically competitive producer who uses labor of a particular type. The representative household supplies all types of labor. The preference functions are assumed to be of the form

$$
\tilde{u}\left(C_{t} ; \xi_{t}\right) \equiv \frac{C_{t}^{1-\tilde{\sigma}^{-1}} \bar{C}_{t}^{\tilde{\sigma}^{-1}}}{1-\tilde{\sigma}^{-1}}, \quad \tilde{v}\left(h_{t}(j) ; \xi_{t}\right) \equiv \frac{\lambda}{1+\nu} h_{t}(j)^{1+\nu} \bar{h}_{t}^{-\nu}
$$

where $\tilde{\sigma}, \nu>0$, and $\left\{\bar{C}_{t}, \bar{h}_{t}\right\}$ are exogenous disturbance processes included in the vector $\xi_{t}$, which is assumed to be bounded up to some date $T$, and constant for all $t>T$.

Each specialized good is produced according to the production function

$$
y_{t}(i)=A_{t} h_{t}(i)^{1 / \phi}
$$

where $A_{t}$ is an exogenously varying technology factor (also included in the vector $\xi_{t}$ ), and $\phi>1$. Implicitly, we consider other factors of production such as the capital stock as being constant or exogenously varying. Aggregate output $Y_{t}$ in turn relates to consumption according to

$$
Y_{t}=C_{t}+G_{t}
$$

where $G_{t}$ denotes exogenous government demand for the composite good and is also included in the vector $\xi_{t}$.

As Benigno and Woodford (2005) show, the utility of the representative household, which is also the policymaker's welfare objective function, can be expressed in the form (1.1) where the period $t$ utility is

$$
\begin{equation*}
\pi\left(y_{t}, \xi_{t}\right)=U\left(Y_{t}, \Delta_{t} ; \xi_{t}\right) \equiv u\left(Y_{t} ; \xi_{t}\right)-v\left(Y_{t} ; \xi_{t}\right) \Delta_{t} \tag{E.1}
\end{equation*}
$$

where

$$
u\left(Y_{t}, \xi_{t}\right) \equiv \tilde{u}\left(Y_{t}-G_{t}, \xi_{t}\right), \quad v\left(Y_{t} ; \xi_{t}\right) \equiv \tilde{v}\left(\left(Y_{t} / A_{t}\right)^{\phi} ; \xi_{t}\right)
$$

express utility as functions of aggregate output and have the properties $\sigma^{-1} \equiv-u_{Y Y} \bar{Y} / u_{Y}=$ $\tilde{\sigma}^{-1} \bar{Y} / \bar{C}>0, \omega \equiv v_{Y Y} \bar{Y} / v_{Y}=v_{Y} \bar{Y} / v-1>0$, and

$$
\Delta_{t} \equiv \int_{0}^{1}\left(\frac{p_{t}(i)}{P_{t}}\right)^{-\theta(1+\omega)} d i \geq 1
$$

is measure of price dispersion at date $t$, where $p_{t}(i)$ denotes the price of individual good $i$ and $P_{t} \equiv\left[\int_{0}^{1} p_{t}(i)^{1-\theta}\right]^{1 /(1-\theta)}$ is a Dixit-Stiglitz price index.

The producers are wage takers on the labor market and choose their prices to maximize the present discounted value of future after-tax nominal profits. As in Calvo's (1983) model of staggered pricing, we assume that producers fix the prices of their goods for a random interval of time, and that a constant fraction $\alpha \in[0,1)$ of prices remain unchanged in any given period. As shown in Benigno and Woodford (2005), since all suppliers that revise their prices in period $t$ face the same problem, they all choose the same new price $p_{t}^{*}$ satisfying the first-order condition

$$
\begin{equation*}
\frac{p_{t}^{*}}{P_{t}}=\left(\frac{K_{t}}{H_{t}}\right)^{\frac{1}{1+\omega \theta}} \tag{E.2}
\end{equation*}
$$

where

$$
\begin{aligned}
H_{t} & \equiv E_{t}\left[\sum_{T=t}^{\infty}(\alpha \beta)^{T-t}\left(1-\varsigma_{T}\right) \cdot u_{Y}\left(Y_{T}, \xi_{T}\right) \cdot Y_{T} \cdot\left(P_{T} / P_{t}\right)^{\theta-1}\right] \\
K_{t} & \equiv E_{t}\left[\sum_{T=t}^{\infty}(\alpha \beta)^{T-t} \frac{\theta \mu_{T}^{w}}{\theta-1} \cdot v_{Y}\left(Y_{T}, \xi_{T}\right) \cdot Y_{T} \cdot\left(P_{T} / P_{t}\right)^{\theta(1+\omega)}\right]
\end{aligned}
$$

and $\varsigma_{t} \in[0,1)$ is an exogenous tax rate on sales revenues and $\mu_{t}^{w} \geq 1$ is an exogenous markup factor on the labor market.

The price index in turn evolves according to a law of motion

$$
\begin{equation*}
P_{t}=\left[(1-\alpha) p_{t}^{* 1-\theta}+\alpha P_{t-1}^{1-\theta}\right]^{1 /(1-\theta)} . \tag{E.3}
\end{equation*}
$$

Combining (E.2) with (E.3) yields

$$
\begin{equation*}
\frac{1-\alpha \Pi_{t}^{\theta-1}}{1-\alpha}=\left(\frac{H_{t}}{K_{t}}\right)^{\frac{\theta-1}{1+\omega \theta}} \tag{E.4}
\end{equation*}
$$

where $\Pi_{t} \equiv P_{t} / P_{t-1}$ is the gross inflation rate. This expression is a short-run aggregate supply relation between inflation and output, given the current disturbances $\xi_{t}$ and expected future inflation, output, and disturbances.

Using again (E.3), we can also obtain an expression for the evolution of the measure of price dispersion

$$
\begin{equation*}
\Delta_{t}=\alpha \Delta_{t-1} \Pi_{t}^{\theta(1+\omega)}+(1-\alpha)\left(\frac{1-\alpha \Pi_{t}^{\theta-1}}{1-\alpha}\right)^{\frac{\theta(1+\omega)}{\theta-1}} \tag{E.5}
\end{equation*}
$$

We assume that the government imposes lump-sum taxes on households so as to guarantee its intertemporal solvency regardless of monetary policy actions. In addition, we abstract from monetary frictions that would generate a demand for money, and assume that the policymaker can control the riskless short-term nominal interest rate and that the lower bound on nominal interest rates never binds. While the optimal intertemporal allocation of households' expenditures determines period- $t$ output as a function of expectations of future output, inflation and the nominal interest, this doesn't constitute a constraint on the policy problem as the central bank can always choose a nominal interest rate that satisfies this equation. As a result, the only relevant constraints facing the policymaker are given by (E.4) and (E.5). While the former prevents the central bank from simultaneously stabilizing inflation and output, the latter determines the evolution of the price dispersion, i.e., a key source of welfare losses.

It will be convenient to rewrite the expressions for $H_{t}, K_{t}$ in recursive form as in (4.2), (4.3), and to use (E.4) to substitute for the variable $\Pi_{t}$ from the system. The resulting restrictions (E.5), (4.2)-(4.3) can then be expressed as in (1.2)-(1.3) where

$$
\begin{align*}
F\left(y_{t}, \xi_{t} ; y_{t-1}\right) & \equiv-\Delta_{t}+\check{f}\left(\Delta_{t-1}, Z_{t}\right)  \tag{E.6}\\
g\left(y_{t}, \xi_{t} ; y_{t+1}\right) & \equiv \check{g}\left(Y_{t}, \xi_{t}\right)-Z_{t}+\alpha \beta \Phi\left(Z_{t+1}\right) \tag{E.7}
\end{align*}
$$

where

$$
\begin{aligned}
\check{f}\left(\Delta_{t-1}, Z_{t}\right) & \equiv \alpha \Delta_{t-1}\left[\frac{1}{\alpha}-\frac{1-\alpha}{\alpha}\left(\frac{H_{t}}{K_{t}}\right)^{\frac{\theta-1}{1+\omega \theta}}\right]^{\frac{\theta(1+\omega)}{\theta-1}}+(1-\alpha)\left(\frac{H_{t}}{K_{t}}\right)^{\frac{\theta(1+\omega)}{1+\omega \theta}} \\
\check{g}\left(Y_{t}, \xi_{t}\right) & =\left[\begin{array}{c}
\left(1-\varsigma_{t}\right) u_{Y}\left(Y_{t}, \xi_{t}\right) Y_{t} \\
\frac{\theta \mu^{w}}{\theta-1} v_{Y}\left(Y_{t}, \xi_{t}\right) Y_{t}
\end{array}\right], \quad \Phi\left(Z_{t}\right)=\left[\begin{array}{c}
\left(\frac{1}{\alpha}-\frac{1-\alpha}{\alpha}\left(\frac{H_{t}}{K_{t}}\right)^{\frac{\theta-1}{1+\omega \theta}}\right) H_{t} \\
\left(\frac{1}{\alpha}-\frac{1-\alpha}{\alpha}\left(\frac{H_{t}}{K_{t}}\right)^{\frac{\theta-1}{1+\omega \theta}}\right)^{\frac{\theta(1+\omega)}{\theta-1}} K_{t}
\end{array}\right] .
\end{aligned}
$$

and the vector of endogenous variables is given by $y_{t} \equiv\left[Y_{t}, \Delta_{t}, H_{t}, K_{t}\right]^{\prime}$, while $Z_{t} \equiv\left[H_{t}, K_{t}\right]^{\prime}$ is a subset of the endogenous variables.

## E. 2 Steady State

We now show that an optimal steady state exists in which the inflation rate is zero $(\bar{\Pi}=1)$. The optimal steady state is described by constant vectors $(\bar{y}, \bar{\varphi}, \bar{\xi})$ satisfying (1.10)-(1.12), or, in this model,

$$
\begin{align*}
-\bar{\Delta}+\check{f}(\bar{\Delta}, \bar{Z}) & =0  \tag{E.8}\\
\check{g}(\bar{Y}, \bar{\xi})-\bar{Z}+\alpha \beta \Phi(\bar{Z}) & =0  \tag{E.9}\\
U_{Y}(\bar{Y}, \bar{\Delta} ; \bar{\xi})+\check{g}_{Y}(\bar{Y}, \bar{\xi})^{\prime} \bar{\Theta} & =0  \tag{E.10}\\
U_{\Delta}(\bar{Y}, \bar{\Delta} ; \bar{\xi})+\left(\beta \check{f}_{\Delta}(\bar{\Delta}, \bar{Z})-1\right) \bar{\theta} & =0  \tag{E.11}\\
\check{f}_{Z}(\bar{\Delta}, \bar{Z})^{\prime} \bar{\theta}+\left(\alpha \Phi_{Z}(\bar{Z})^{\prime}-I_{2}\right) \bar{\Theta} & =0, \tag{E.12}
\end{align*}
$$

where $\bar{Z}=[\bar{H}, \bar{K}]^{\prime}$ together with the steady-state versions of equation (E.4). We proceed by conjecturing that the solution involves $\bar{\Pi}=1$, and showing that a solution can be constructed that satisfies all of the equations just listed.

We first observe that (E.4) implies that a steady state with $\bar{\Pi}=1$ must satisfy $\bar{H}=\bar{K}$. Given this, (E.8) requires that $\bar{\Delta}=1$, so that there is zero price dispersion. Condition (E.9) holds as well if and only if $\bar{Y}$ is the output level implicitly defined by

$$
\begin{equation*}
(1-\varsigma) u_{Y}(\bar{Y}, \bar{\xi})=\frac{\theta \bar{\mu}^{w}}{\theta-1} v_{Y}(\bar{Y}, \bar{\xi}) \tag{E.13}
\end{equation*}
$$

and $\bar{H}=\bar{K}=(1-\alpha \beta)^{-1} \frac{\theta \mu^{w}}{\theta-1} v_{Y}(\bar{Y}, \bar{\xi}) \bar{Y}$.
Because $\check{f}_{Z}(1, \bar{Z})=0$, (E.12) reduces to the eigenvector equation

$$
\begin{equation*}
\left(\alpha \Phi_{Z}(\bar{Z})^{\prime}-I_{2}\right) \bar{\Theta}=0 \tag{E.14}
\end{equation*}
$$

where

$$
\Phi_{Z}(\bar{Z})=\left[\begin{array}{cc}
\frac{1}{\alpha}+\frac{\theta(\alpha-1)(\omega+1)}{\alpha(\theta \omega+1)} & -\frac{(\theta-1)(\alpha-1)}{\alpha(\theta+1)} \\
\frac{\theta(\alpha-1)(\omega+1)}{\alpha(\theta \omega+1)} & 1-\frac{\theta(\alpha-1)(\omega+1)}{\alpha(\theta \omega+1)}
\end{array}\right] .
$$

Since $\Phi_{Z}(\bar{Z})^{\prime}$ has an eigenvector $[-1,1]^{\prime}$, with eigenvalue $1 / \alpha$, (E.14) is satisfied if and only if $\bar{\Theta}_{2}=-\bar{\Theta}_{1}$.

Conditions (E.10)-(E.11) in turn allow us to determine the values $\bar{\theta}, \bar{\Theta}$. Given that $\check{f}_{\Delta}(\bar{\Delta}, \bar{Z})=\alpha$ and

$$
\check{g}_{Y}(\bar{Y}, \bar{\xi})=\left[\begin{array}{c}
(1-\varsigma)\left(u_{Y Y} \bar{Y}+u_{Y}\right) \\
\frac{\theta \omega^{w}}{\theta-1}\left(v_{Y Y} \bar{Y}+v_{Y}\right)
\end{array}\right]
$$

we obtain

$$
\bar{\theta}=-\frac{v(\bar{Y}, \bar{\xi})}{1-\alpha \beta}, \quad \bar{\Theta}_{1}=-\bar{\Theta}_{2}=\frac{\Psi}{(1-\varsigma)\left(\omega+\sigma^{-1}\right)}
$$

where

$$
\Psi \equiv 1-\frac{\theta-1}{\theta} \frac{1-\varsigma}{\mu^{w}}
$$

is a measure of the degree of inefficiency of the steady-state output level $\bar{Y}$. Here we use the fact that $U_{Y}=u_{Y}-v_{Y}=\Psi u_{Y}$ and $u_{Y Y}-v_{Y Y}=-\frac{u_{Y Y}}{Y}\left(\omega(1-\Psi)+\sigma^{-1}\right)$. (The first of these equations explains our interpretation of $\Psi$ as a measure of the degree of inefficiency: a positive value of $\Psi$ indicates that utility would be increased by raising $\bar{Y}$, maintaining zero price dispersion and hence an equal level of production of each of the differentiated goods.)

## E. 3 Canonical Decomposition of $\bar{A}-\mu \bar{I}$

Evaluating the derivatives of $F()$ and $g()$ at the steady state, we can then construct the matrices $\bar{A}$ and $\bar{I}$ :

$$
\begin{aligned}
& \bar{A} \equiv\left[\begin{array}{c}
\beta D_{3} F \\
D_{1} g
\end{array}\right]=\left[\begin{array}{cccc}
0 & \alpha \beta & 0 & 0 \\
a_{21} & 0 & -1 & 0 \\
a_{31} & 0 & 0 & -1
\end{array}\right] \\
& \bar{I} \equiv\left[\begin{array}{c}
-\beta D_{1} F \\
-D_{3} g
\end{array}\right]=\left[\begin{array}{cccc}
0 & \beta & 0 \\
0 & 0 & -\alpha \beta+\frac{\beta(1-\alpha)(\theta-1)}{1+\omega \theta} & -\frac{\beta(1-\alpha)(\theta-1)}{1+\omega \theta} \\
0 & 0 & \frac{\beta(1-\alpha) \theta(1+\omega)}{1+\omega \theta} & -\alpha \beta-\frac{\beta(1-\alpha) \theta(1+\omega)}{1+\omega \theta}
\end{array}\right],
\end{aligned}
$$

where

$$
a_{21}=(1-\varsigma) u_{Y}\left(1-\sigma^{-1}\right), \quad a_{31}=(1-\varsigma) u_{Y}(1+\omega)
$$

The matrix pencil $\bar{A}-\mu \bar{I}$ pencil satisfies Assumption 2. We first determine the minimal degree associated with the matrix pencil $\bar{A}-\mu \bar{I}$. Following Gantmacher (1959, chap. 12, p. 30), the minimal degree of the matrix pencil $\bar{A}-\mu \bar{I}$ is the least value of the index $l$ for which the rank of the matrix

$$
M_{l}=\overbrace{\left[\begin{array}{cccc}
\bar{A} & 0 & \cdots & 0 \\
-\bar{I} & \bar{A} & & \vdots \\
0 & -\bar{I} & \ddots & 0 \\
\vdots & & \ddots & \bar{A} \\
0 & 0 & \cdots & -\bar{I}
\end{array}\right]}^{l+1}
$$

satisfies

$$
\operatorname{rank}\left(M_{l}\right)<(l+1) m .
$$

We observe that $\operatorname{rank}\left(M_{0}\right)=4=(0+1) 4, \operatorname{rank}\left(M_{1}\right)=8=(1+1) 4$, but that $\operatorname{rank}\left(M_{2}\right)=$ $11<(2+1) 4=12$. The minimal degree of the matrix pencil $\bar{A}-\mu \bar{I}$ is therefore $\tilde{n}=2$, so that $q=n-\tilde{n}=1$.

It follows from Lemma 3 that there exist nonsingular matrices $P$ and $Q$ of dimensions $3 \times 3$ and $4 \times 4$ respectively that satisfy

$$
(\bar{A}-\mu \bar{I})^{\prime}=Q\left[\begin{array}{ccc}
\mu & 0 & 0 \\
1 & \mu & 0 \\
0 & 1 & 0 \\
0 & 0 & B_{2}^{\prime}-\mu J_{2}^{\prime}
\end{array}\right] P
$$

These matrices are given by

$$
Q^{-1}=\left[\begin{array}{cccc}
1 & 0 & a_{21} & a_{31} \\
\beta(1+\alpha) & 0 & q_{23} & q_{24} \\
\alpha \beta^{2} & 0 & 0 & 0 \\
0 & 1 & 0 & 0
\end{array}\right], \quad P=\left[\begin{array}{ccc}
0 & \beta a_{21}(\alpha+1)-q_{23} & \beta a_{31}(\alpha+1)-q_{24} \\
0 & \alpha \beta^{2} a_{21} & \alpha \beta^{2} a_{31} \\
\beta & 0 & 0
\end{array}\right]
$$

where $\operatorname{rank}\left(Q^{-1}\right)=4$ and $\operatorname{rank}(P)=3$, and $q_{23}, q_{24}$ satisfy $q_{23}-q_{24}=\alpha \beta\left(a_{21}-a_{31}\right)$. We also have $B_{2}=\alpha$ and $J_{2}=1$.

## E. 4 Target Variables

When evaluating the second derivatives of $F(), g()$ and the objective function $\pi()$ at the optimal steady state, the target variables $\tau_{t}$ and the target values $\tau_{t}^{*}$ are given by

$$
\begin{aligned}
\tau_{t} & =-\left[S \tilde{y}_{t}+R \tilde{y}_{t-1}+\beta R^{\prime} E_{t} \tilde{y}_{t+1}\right] \\
\tau_{t}^{*} & =\left\{D_{2}\left[\left(D_{1} \pi\right)^{\prime}\right]+\left(\bar{\Theta}^{\prime} \otimes I_{4}\right) D_{2}\left[\left(D_{1} g\right)^{\prime}\right]\right\} \tilde{\xi}_{t},
\end{aligned}
$$

where the matrices $S$ and $R$ reduce to

$$
\begin{aligned}
S & \equiv D_{1}\left[\left(D_{1} \pi\right)^{\prime}\right]+\bar{\theta} D_{1}\left[\left(D_{1} F\right)^{\prime}\right]+\left(\bar{\Theta}^{\prime} \otimes I_{4}\right)\left\{D_{1}\left[\left(D_{1} g\right)^{\prime}\right]+\beta^{-1} D_{3}\left[\left(D_{3} g\right)^{\prime}\right]\right\} \\
& =\left[\begin{array}{ccc}
U_{Y Y}+\bar{\Theta}^{\prime} \check{g}_{Y Y} & U_{Y \Delta} & 0_{1 \times 2} \\
U_{\Delta Y} & 0 & 0_{1 \times 2} \\
0 & 0 & \bar{\theta} \check{f}_{Z Z}(\bar{\Delta}, \bar{Z})+\left(\bar{\Theta}^{\prime} \otimes I_{2}\right) \alpha \Phi_{Z Z}
\end{array}\right], \\
R & \equiv \bar{\theta} D_{3}\left[\left(D_{1} F\right)^{\prime}\right]=\left[\begin{array}{ccc}
0_{2 \times 1} & 0_{2 \times 1} & 0_{2 \times 2} \\
0_{2 \times 1} & \bar{\theta} \check{f}_{Z \Delta}(\bar{\Delta}, \bar{Z}) & 0_{2 \times 2}
\end{array}\right] .
\end{aligned}
$$

Here, we use properties of the second derivatives of $F()$ and $g(): D_{3}\left[\left(D_{3} F\right)^{\prime}\right]=0, D_{2}\left[\left(D_{1} F\right)^{\prime}\right]=$ $0, D_{2}\left[\left(D_{3} F\right)^{\prime}\right]=0, D_{1}\left[\left(D_{3} g\right)^{\prime}\right]=0, D_{2}\left[\left(D_{3} g\right)^{\prime}\right]=0$, and

$$
\left(\bar{\Theta}^{\prime} \otimes I_{4}\right) D_{1}\left[\left(D_{1} g\right)^{\prime}\right]=\left[\begin{array}{cc}
\bar{\Theta}^{\prime} \check{g}_{Y Y} & 0 \\
0 & 0_{3 \times 3}
\end{array}\right], \quad\left(\bar{\Theta}^{\prime} \otimes I_{4}\right) D_{2}\left[\left(D_{1} g\right)^{\prime}\right]=\left[\begin{array}{c}
\bar{\Theta}^{\prime} \check{g}_{Y \xi} \\
0_{3 \times n_{\xi}}
\end{array}\right],
$$

where $\check{g}_{Y Y}=\partial \check{g}_{Y}(\bar{Y}, \bar{\xi}) / \partial Y, \check{g}_{Y \xi}=\partial \check{g}_{Y}(\bar{Y}, \bar{\xi}) / \partial \xi^{\prime}$. Similarly, we use

$$
\left(\bar{\Theta}^{\prime} \otimes I_{4}\right) D_{3}\left[\left(D_{3} g\right)^{\prime}\right]=\left[\begin{array}{cc}
0_{2 \times 2} & 0_{2 \times 2} \\
0_{2 \times 2} & \left(\bar{\Theta}^{\prime} \otimes I_{2}\right) \alpha \beta \Phi_{Z Z}
\end{array}\right]
$$

where $\Phi_{Z Z}=\partial v e c\left(\Phi_{Z}(\bar{Z})^{\prime}\right) / \partial Z^{\prime}$. Moreover, since

$$
\left\{D_{2}\left[\left(D_{1} \pi\right)^{\prime}\right]+\left(\bar{\Theta}^{\prime} \otimes I_{4}\right) D_{2}\left[\left(D_{1} g\right)^{\prime}\right]\right\}=\left[\begin{array}{c}
U_{Y \xi}+\bar{\Theta}^{\prime} \check{g}_{Y \xi} \\
U_{\Delta \xi} \\
0_{2 \times 1}
\end{array}\right]
$$

we can express the target gaps as

$$
\tau_{t}-\tau_{t}^{*}=-\left[\begin{array}{c}
\left(U_{Y Y}+\bar{\Theta}^{\prime} \check{g}_{Y Y}\right) \tilde{Y}_{t}+U_{Y \Delta} \tilde{\Delta}_{t}+\left(U_{Y \xi}+\bar{\Theta}^{\prime} \check{g}_{Y \xi}\right) \tilde{\xi}_{t} \\
U_{\Delta Y} \tilde{Y}_{t}+\beta \tilde{f}_{Z \Delta}(\bar{\Delta}, \bar{Z})^{\prime} \bar{\theta} E_{t} \tilde{Z}_{t+1}+U_{\Delta \xi} \tilde{\xi}_{t} \\
\left(\bar{\theta} \check{f}_{Z Z}(\bar{\Delta}, \bar{Z})+\left(\bar{\Theta}^{\prime} \otimes I_{2}\right) \alpha \Phi_{Z Z}\right) \tilde{Z}_{t}+\bar{\theta} \check{f}_{Z \Delta}(\bar{\Delta}, \bar{Z}) \tilde{\Delta}_{t-1}
\end{array}\right]
$$

Linearizing the relationship implicitly defining $Y_{t}^{*}$, (4.6), we have

$$
U_{Y Y} \tilde{Y}_{t}^{*}+U_{Y \xi} \tilde{\xi}_{t}+\bar{\Theta}^{\prime}\left(\check{g}_{Y Y} \tilde{Y}_{t}^{*}+\check{g}_{Y \xi} \tilde{\xi}_{t}\right)=0
$$

where $\tilde{Y}_{t}^{*}=Y_{t}^{*}-\bar{Y}$. This allows us to rewrite the first element of $\tau_{t}-\tau_{t}^{*}$ as

$$
\left(U_{Y Y}+\bar{\Theta}^{\prime} \check{g}_{Y Y}\right) \bar{Y} x_{t}+U_{Y \Delta} \tilde{\Delta}_{t},
$$

where $x_{t}=\log \left(Y_{t} / Y_{t}^{*}\right)=\left(\tilde{Y}_{t}-\tilde{Y}_{t}^{*}\right) / \bar{Y}+\mathcal{O}\left(\epsilon^{2}\right)$ measures the welfare-relevant output gap. The target gaps $\tau_{t}-\tau_{t}^{*}$ can further be simplified, by noting that a linear approximation of (E.4) yields

$$
\pi_{t}=-\frac{1-\alpha}{(1+\omega \theta) \alpha \bar{K}}\left(\tilde{H}_{t}-\tilde{K}_{t}\right)
$$

where $\pi_{t} \equiv \log \Pi_{t}$. Using this and the fact that $\bar{\Theta}_{1}=-\bar{\Theta}_{2}$, we obtain

$$
\begin{aligned}
\check{f}_{Z Z}(\bar{\Delta}, \bar{Z}) \tilde{Z}_{t} & =-\frac{1}{\bar{K}} \theta(1+\omega)\left[\begin{array}{c}
1 \\
-1
\end{array}\right] \pi_{t} \\
\left(\bar{\Theta}^{\prime} \otimes I_{2}\right) \alpha \Phi_{Z Z} \tilde{Z}_{t} & =\bar{\Theta}_{1} \theta(\omega+1)\left[\begin{array}{c}
1 \\
-1
\end{array}\right] \pi_{t},
\end{aligned}
$$

so that the target gaps can be rewritten as

$$
\tau_{t}-\tau_{t}^{*}=-\left[\begin{array}{c}
\left(U_{Y Y}+\bar{\Theta}^{\prime} \check{g}_{Y Y}\right) \bar{Y} x_{t}+U_{Y \Delta} \tilde{\Delta}_{t}  \tag{E.15}\\
U_{\Delta Y} \tilde{Y}_{t}+\beta \check{f}_{Z \Delta}(\bar{\Delta}, \bar{Z})^{\prime} \bar{\theta} E_{t} \tilde{Z}_{t+1}+U_{\Delta \xi} \tilde{\xi}_{t} \\
\theta(\omega+1)\left(\left(\bar{\Theta}_{1}-\frac{\bar{\theta}}{K}\right) \pi_{t}-\frac{\bar{\theta}}{\bar{K}} \frac{(1-\alpha)}{(1++\omega)} \tilde{\Delta}_{t-1}\right) \\
-\theta(\omega+1)\left(\left(\bar{\Theta}_{1}-\frac{\bar{\theta}}{\bar{K}}\right) \pi_{t}-\frac{\bar{\theta}}{\bar{K}} \frac{(1-\alpha)}{(1+\theta \omega)} \tilde{\Delta}_{t-1}\right)
\end{array}\right] .
$$

## E. 5 Optimal Target Criterion

Since $k_{1}=0$, the general optimal target criterion (3.20) reduces to

$$
0=z_{t} \equiv\left[\begin{array}{ccc}
1 & -\beta^{-1} L & \beta^{-2} L^{2}
\end{array}\right] \hat{\tau}_{1 t}
$$

where

$$
\hat{\tau}_{1 t}=\left[\begin{array}{ll}
I_{m-q} & 0
\end{array}\right] Q^{-1}\left(\tau_{t}-\tau_{t}^{*}\right)
$$

Combining this with (E.15), we obtain

$$
\left.\left.\left.\begin{array}{rl}
0= & {\left[\begin{array}{lll}
1-(1+\alpha) L+\alpha L^{2} & 0 & a_{21}-q_{23} \beta^{-1} L
\end{array} a_{31}-q_{24} \beta^{-1} L\right.}
\end{array}\right]\left(\tau_{t}-\tau_{t}^{*}\right)\right] \text { = }(1-\alpha L)(1-L)\left(\left(U_{Y Y}+\bar{\Theta}^{\prime} \check{g}_{Y Y}\right) \bar{Y} x_{t}+U_{Y \Delta} \tilde{\Delta}_{t}\right)\right] .
$$

using $\left(q_{23}-q_{24}\right)=\alpha \beta\left(a_{21}-a_{31}\right)$ to obtain the last equality.
Noting furthermore that a first-order approximation of (E.6) around the optimal steady state yields

$$
0=-\tilde{\Delta}_{t}+\check{f}_{\Delta}(\bar{\Delta}, \bar{Z}) \tilde{\Delta}_{t-1}+\check{f}_{Z}(\bar{\Delta}, \bar{Z}) \tilde{Z}_{t}
$$

or

$$
(1-\alpha L) \tilde{\Delta}_{t}=0
$$

for all $t$, we observe that the variable $\tilde{\Delta}_{t}$ drops out from the target criterion, so that (E.16) reduces to

$$
0=(1-\alpha L)\left[\left(U_{Y Y}+\bar{\Theta}^{\prime} \check{g}_{Y Y}\right) \bar{Y}(1-L) x_{t}+\left(a_{21}-a_{31}\right) \theta(\omega+1)\left(\bar{\Theta}_{1}-\frac{\bar{\theta}}{\bar{K}}\right) \pi_{t}\right]
$$

or equivalently

$$
0=(1-\alpha L)\left(\pi_{t}+\phi(1-L) x_{t}\right)
$$

where the weight $\phi$ on changes in output gap fluctuations is given by

$$
\begin{equation*}
\phi=\frac{\left(U_{Y Y}+\bar{\Theta}^{\prime} \check{g}_{Y Y}\right) \bar{Y}}{\left(a_{21}-a_{31}\right) \theta(\omega+1)\left(\bar{\Theta}_{1}-\frac{\bar{\theta}}{K}\right)}=\frac{\omega+\sigma^{-1}+\Psi\left(1-\sigma^{-1}\right)-\frac{\Psi \sigma^{-1}(\bar{Y} / \bar{C}-1)}{\omega+\sigma^{-1}}}{\theta\left(\omega+\sigma^{-1}+\Psi\left(1-\sigma^{-1}\right)\right)} . \tag{E.17}
\end{equation*}
$$

To obtain the second equality in (E.17), we use (E.13) and the properties of the preference functions $\frac{u_{Y Y Y} \bar{Y}^{2}}{u_{Y}}=\sigma^{-1}\left(\sigma^{-1}+\bar{Y} / \bar{C}\right)$ and $\frac{v_{Y Y Y} \bar{Y}^{2}}{v_{Y}}=\omega(\omega-1)$. We thus obtain the representation (4.5) for the optimal target criterion, as stated in the text.


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[^1]:    ${ }^{23}$ This demonstration that the eigenvalues come in "reciprocal pairs" extends to our environment a standard result in the theory of linear-quadratic optimal control (e.g., Hansen and Sargent, 2010, chap. 8).

[^2]:    ${ }^{24}$ Note that convergence in the exact nonlinear dynamics only implies that the sequence must not explode in the linearized dynamics, since the rate of convergence might asymptotically decrease to zero.

[^3]:    ${ }^{25}$ Note that $\left(I+c J_{0}-\mu J_{0}\right)\left(I+c J_{0}\right)^{-1}=I-\mu J_{0}\left(I+c J_{0}\right)^{-1}=I-\mu\left(J_{0}^{-1}\right)^{-1}\left(I+c J_{0}\right)^{-1}=I-$ $\mu\left(\left(I+c J_{0}\right) J_{0}^{-1}\right)^{-1}=I-\mu\left(J_{0}^{-1}+c I\right)^{-1}=I-\mu\left(J_{0}^{-1}\left(I+c J_{0}\right)\right)^{-1}=I-\mu\left(I+c J_{0}\right)^{-1} J_{0}$.

