# COMPUTING $\exp (-\tau \boldsymbol{A}) b$ WITH LAGUERRE POLYNOMIALS* 

BERNARD N. SHEEHAN ${ }^{\dagger}$, YOUSEF SAAD ${ }^{\ddagger}$, AND ROGER B. SIDJE§


#### Abstract

This paper discusses a method based on Laguerre polynomials combined with a Filtered Conjugate Residual (FCR) framework to compute the product of the exponential of a matrix by a vector. The method implicitly uses an expansion of the exponential function in a series of orthogonal Laguerre polynomials, much like existing methods based on Chebyshev polynomials do. Owing to the fact that orthogonal polynomials satisfy a three-term recurrence, what these series expansion methods offer over other approaches such as Krylov subspace methods lies in the elimination of inner products and the economy in storage since there is no need to compute and keep a set of basis vectors. Compared with Chebyshev polynomials that are orthogonal within a restricted interval and need estimates of the outermost eigenvalues, Laguerre polynomials offer the added feature that they are orthogonal on the half real line, alleviating therefore the need to estimate eigenvalues.


Key words. Conjugate residual, filtered conjugate residual, polynomial filtering, exponential propagation, orthogonal polynomials

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1. Introduction. The problem of calculating expressions of the form $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$, where $\boldsymbol{A} \in \mathbb{R}^{n \times n}$ is a non-negative definite matrix and $\boldsymbol{b} \in \mathbb{R}^{n}$ an arbitrary vector, occurs frequently in applications. The problem is equivalent, for example, to that of finding the solution to the system of ordinary differential equations

$$
\begin{equation*}
\dot{\boldsymbol{y}}=-\boldsymbol{A} \boldsymbol{y}, \quad \boldsymbol{y}(0)=\boldsymbol{b} \tag{1.1}
\end{equation*}
$$

at time $\tau$. Equation (1.1), in turn, arises out of finite difference or finite element discretizations of thermal problems ( $\boldsymbol{A}$ symmetric) or convection-diffusion and vibration problems ( $\boldsymbol{A}$ non-symmetric). Similarly, very stiff systems of the form (1.1) arise in predicting the time evolution of electrical circuits. Another occurrence of (1.1) is in computing the transient solution of Markov chains [22].

The calculation of a matrix exponential times a vector can be a treacherous task; see [17] for a survey of potential difficulties. Many methods have been proposed [4, 6, 9, 10, 14, 16]. For very large, sparse matrices, perhaps the preferred method $[7,8,11-13,18,21]$ is to use the Lanczos or Arnold procedure to obtain a matrix $\boldsymbol{V}_{m} \in \mathbb{R}^{n \times m}$ whose columns span the Krylov subspace span $\left\{\boldsymbol{b}, \boldsymbol{A} \boldsymbol{b}, \cdots, \boldsymbol{A}^{m-1} \boldsymbol{b}\right\}$, and then to write

$$
\begin{equation*}
\exp (-\tau \boldsymbol{A}) \boldsymbol{b} \approx \boldsymbol{V}_{m} \exp \left(-\tau \boldsymbol{H}_{m}\right) \beta \boldsymbol{e}_{1}, \tag{1.2}
\end{equation*}
$$

where $\boldsymbol{H}_{m}$ is the tridiagonal or Hessenberg matrix resulting from the Lanczos or Arnoldi process and $\beta=\|\boldsymbol{b}\|_{2}$. Since $\boldsymbol{H}_{m}$ will normally be much smaller than $\boldsymbol{A}$, dense matrix methods such as Padé approximations to $\exp (t)$ can be used to evaluate $\exp \left(-\tau \boldsymbol{H}_{m}\right)$.

Bergamaschi et al. have reported experiments that use expansions of the exponential function in Chebyshev polynomials [2,3]. The paper convincingly argued that Chebyshevbased methods can give the same accuracy as Krylov techniques using the same polynomial degree, but that they can be much less expensive as they require no inner products. The fact

[^0]that this technique appears to be competitive with (1.2) prompted us to ask whether other choices of orthogonal polynomials might be used to good effect.

This work was initially motivated by an intriguing question. The use of Chebyshev polynomials requires some prior knowledge of an interval $[a, b]$ which contains the spectrum of the matrix [2, 3]. In contrast, polynomials that are orthogonal on the half real line (e.g., Laguerre) or the whole line (Hermite) would normally require no bounds. This is an important practical advantage. For example, one can ask the question: Is it possible to employ a Laguerre series expansion to obtain an inexpensive exponential propagator which bypasses completely both inner products and eigenvalue estimates? We will show that if used with care, Laguerre expansions combined with a good implementation of the orthogonal expansion can be quite effective. On the other hand, they are as inexpensive as methods based on Chebyshev expansions and do not require accurate eigenvalue estimates. Recall that Chebyshev-based techniques require the knowledge of an interval which is guaranteed to contain all eigenvalues. If the interval fails to contain all eigenvalues the technique will not work. A common remedy is to take a large interval containing all eigenvalues, such as one provided by the Gershgorin theorem, but often this yields poor convergence.

The techniques we examine in this paper are based on the expansion of $e^{-\tau t}$ as a series of Laguerre polynomials. The filtered conjugate residual-like algorithm (FCR) introduced in [20] provides a framework for exploiting least-squares polynomials to solve problems as diverse as regularization in graphics, information retrieval, and in electronic structure calculations. Here, we will use this framework again and show that an FCR-type algorithm can also be usefully applied to exponential propagation. At the same time, we also consider the same framework applied to expansions in Chebyshev polynomials for the purpose of comparison.
2. Computing $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ with orthogonal polynomials. It is possible to approximate $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ by using either rational or polynomial approximations to $\exp (-\tau t)$. Methods based or rational approximations require solving large sparse linear systems of equations and are not considered here. We consider methods based on approximating $\exp (-\tau t)$ by a polynomial $p(t)$. Thus, a polynomial $p_{m}$ of degree $m$ is found which approximates $\exp (-\tau t)$, and $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ is approximated by $p_{m}(\boldsymbol{A}) \boldsymbol{b}$. Note that the matrix $p_{m}(\boldsymbol{A})$ is not computed. Instead, $p_{m}(\boldsymbol{A}) \boldsymbol{b}$ is evaluated by a series of matrix-vector multiplies using $\boldsymbol{A}$.
2.1. Orthogonal expansions for $\exp (-\tau t)$. As background to this strategy, we briefly consider in this section how to expand $\exp (-\tau t)$ in series of generalized Laguerre polynomials and Chebyshev polynomials.

The generalized Laguerre Polynomials $L_{n}^{\alpha}(t), n=0,1,2, \ldots$ and $\alpha>-1$, are orthogonal with respect to the inner product

$$
\begin{equation*}
\langle p, q\rangle_{\alpha}=\int_{0}^{\infty} t^{\alpha} e^{-t} p(t) q(t) d t \tag{2.1}
\end{equation*}
$$

The expansion of the exponential function $e^{-\tau t}$ in terms of Laguerre polynomials is known to be [15, p. 90]

$$
\begin{equation*}
e^{-\tau t}=(\tau+1)^{-\alpha-1} \sum_{n=0}^{\infty}\left(\frac{\tau}{\tau+1}\right)^{n} L_{n}^{\alpha}(t), \quad 0<t<\infty \tag{2.2}
\end{equation*}
$$

By truncating the above summation to only $m$ terms, a polynomial of degree $m$ will be obtained that will approximate $e^{-\tau t}$ over some interval. It is also possible to use expansions in Hermite polynomials, which are orthogonal on the whole real line with respect to the weight $e^{-t^{2}}$, but we will not consider this in this paper.

The Chebyshev polynomials $T_{n}(t), n=0,1,2, \ldots$, are orthogonal over the interval $[-1,1]$ with respect to the inner product

$$
\begin{equation*}
\langle p, q\rangle_{T}=\int_{-1}^{+1} \frac{p(t) q(t) d t}{\sqrt{1-t^{2}}} \tag{2.3}
\end{equation*}
$$

In this case, it is useful to give the expansion of the exponential that is fitted in a general interval $[a, b]$, rather than just the interval $[-1,1]$. The expansion of $e^{-\tau t}$ in terms of Chebyshev polynomials over $[a, b]$ is given by [1, Section 9.6]

$$
\begin{align*}
e^{-\tau t} & =\sum_{k=0}^{\infty} a_{k} T_{k}\left(\frac{t-l_{2}}{l_{1}}\right), \quad a<t<b \quad \text { with }  \tag{2.4}\\
a_{0} & =e^{-\tau l_{2}} I_{0}\left(-\tau l_{1}\right), \quad a_{k}=2 e^{-\tau l_{2}} I_{k}\left(-\tau l_{1}\right), \quad k>1 . \tag{2.5}
\end{align*}
$$

Here, $I_{k}(t)$ is the modified Bessel function of the first kind, and $l_{1}=(b-a) / 2$ and $l_{2}=(a+b) / 2$ are the semi-width and the midpoint, respectively, of the interval over which the approximation is desired. For further information on orthogonal polynomials, the reader is referred to $[1,15]$.
2.2. Classical use of expansions in orthogonal polynomials. Standard orthogonal polynomials satisfy recurrence relations of the form

$$
\begin{equation*}
\beta_{n+1} P_{n+1}(t)=\left(t-\alpha_{n}\right) P_{n}(t)-\gamma_{n} P_{n-1}(t), n=0,1, \ldots, \tag{2.6}
\end{equation*}
$$

with the convention that for $n=0$ the term $\gamma_{0} p_{-1}$ is zero. Thus, Chebyshev polynomials (of the first kind) satisfy the well-known recurrence $T_{n+1}(t)=2 t T_{n}(t)-T_{n-1}(t)$ starting with $T_{0}(t)=1$ and $T_{1}(t)=t$. Similarly, the three-term recurrence for the generalized Laguerre polynomials is

$$
\begin{aligned}
-(n+1) L_{n+1}^{\alpha}(t) & =(t-\alpha-2 n-1) L_{n}^{\alpha}(t)+(n+\alpha) L_{n-1}^{\alpha}(t), \quad n=1,2, \ldots \\
L_{0}^{\alpha}(t) & =1, \quad L_{1}^{\alpha}(t)=1+\alpha-t
\end{aligned}
$$

These recurrences allow one to easily generate successive members of these orthogonal families. Assuming $e^{-\tau t}$ has the following expansion in terms of orthogonal polynomials $P_{n}(t)$ (see (2.2), and (2.4)-(2.5)),

$$
\begin{equation*}
e^{-\tau t}=\sum_{n=0}^{\infty} c_{n} P_{n}(t) \tag{2.7}
\end{equation*}
$$

the following algorithm can be used to compute an approximation to $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ :
Algorithm 2.1. $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ by Orthogonal Expansions

1. $\boldsymbol{p}_{0}=P_{0}(\boldsymbol{A}) \boldsymbol{b} ; \quad \boldsymbol{p}_{1}=P_{1}(\boldsymbol{A}) \boldsymbol{b}$
2. $\boldsymbol{z}=c_{0} \boldsymbol{p}_{0}+c_{1} \boldsymbol{p}_{1}$
3. For $j=1,2, \ldots$, Do:
4. $\quad \boldsymbol{p}_{j+1}=\left(\boldsymbol{A} \boldsymbol{p}_{j}-\alpha_{j} \boldsymbol{p}_{j}-\gamma_{j} \boldsymbol{p}_{j-1}\right) / \beta_{j+1}$
5. $\boldsymbol{z}=\boldsymbol{z}+c_{j+1} \boldsymbol{p}_{j+1}$
6. if $\left|c_{j+1}\right|\left\|\boldsymbol{p}_{j+1}\right\| \leq \epsilon$, break
7. EndDo

The algorithm uses (2.6) in line 4 and (2.7) in lines 2 and 5. If $\boldsymbol{A}$ is an $n \times n$ matrix with $N z(\boldsymbol{A})$ non-zeros, then Algorithm 2.1 entails a computational cost of about $9 n+2 N z(\boldsymbol{A})$ operations per pass through the for-loop. Two vectors are needed to store $\boldsymbol{p}_{j}$ and $\boldsymbol{p}_{j-1}$. These
are exchanged to the pair $\boldsymbol{p}_{j}, \boldsymbol{p}_{j+1}$ in the next step, which can be carried out in a constant number of operations by redirecting pointers rather than copying data. Another vector is needed to perform the matrix-vector product $\boldsymbol{A} \boldsymbol{p}_{j}$ and another one to store the solution $\boldsymbol{z}$. This brings the total number of vectors to four.
2.3. Filtered conjugate residual-type algorithm. An alternative way of computing $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ using orthogonal polynomials is based on the Filtered Conjugate Residual (FCR) algorithm presented in [20]. The algorithm parallels the usual Conjugate Residual algorithm [19] except that constants $\tilde{\alpha}_{j}$ and $\beta_{j}$ are computed using an inner product $\langle,\rangle_{w}$ in function space rather than by the usual vector inner product.

In order to present the Conjugate Residual method for $\exp (-\tau A) b$ in a way that is similar to the Conjugate Residual method for linear systems, it is necessary to reformulate the problem slightly. We would like to approximate a function $\psi(t)$ by polynomials of the form $t s_{k}(t)$ where $s_{k}$ is a polynomial of degree $k$. In the usual context of linear systems, $\psi(t) \equiv 1$ and $s_{k}(\boldsymbol{A}) \boldsymbol{b}$ is the approximate solution (assuming $\boldsymbol{x}_{0}=\mathbf{0}$ ), while $\boldsymbol{b}-\boldsymbol{A} s_{k}(\boldsymbol{A}) \boldsymbol{b}$ is the residual. For the exponential function we will proceed indirectly by instead approximating $\psi(t)=1-e^{-\tau t}$ by a polynomial of the form $t s_{k}(k)$. Thus, the actual polynomial for $e^{-\tau t}$ is $\rho_{k}(t)=1-t s_{k}(t)$, which is the usual residual polynomial in the context of linear systems. One notable advantage of this approach compared to that of Algorithm 2.1 is that the approximation will be exact at $t=0$. In other words, $\rho_{k}(0)=\exp (-\tau \cdot 0)=1$. This property may or may not be important depending on the situation. For example, when solving ordinary differential equations, this property is vital as it ensures that the underlying integration scheme is at least first-order accurate.

Consider the (functional) inner product

$$
\langle p, q\rangle_{w}=\int_{a}^{b} p(t) q(t) w(t) d t
$$

where $w(t) \geq 0$ is some weight function. This inner product induces a weighted $L_{2}$-norm $\|p\|_{w}=\langle p, p\rangle^{1 / 2}$. The goal is to find the polynomial of the form $t s_{k}(t)$ which is closest to $\psi(t)$ in the sense of this $L_{2}$ norm, i.e., such that

$$
\left\|\psi-t s_{k}(t)\right\|_{w}=\min _{s \in \mathcal{P}_{k}}\|\psi-t s(t)\|_{w} .
$$

The algorithm described in [20] exploits an analogy with what is known when solving linear systems. The Conjugate Residual (CR) algorithms to minimize $\|\boldsymbol{b}-\boldsymbol{A} s(\boldsymbol{A}) \boldsymbol{b}\|_{2}$, is as follows:

Algorithm 2.2. Conjugate Residual Algorithm
0. Compute $\boldsymbol{r}_{0}:=\boldsymbol{b}$ (starts with $\boldsymbol{x}_{0}=\mathbf{0}$ ), $\boldsymbol{p}_{0}:=\boldsymbol{r}_{0}$
2. For $j=0,1, \ldots$, until convergence Do:
3. $\alpha_{j}:=\left\langle\boldsymbol{r}_{j}, \boldsymbol{A} \boldsymbol{r}_{j}\right\rangle /\left\langle\boldsymbol{A} \boldsymbol{p}_{j}, \boldsymbol{A} \boldsymbol{p}_{j}\right\rangle$
4. $\boldsymbol{x}_{j+1}:=\boldsymbol{x}_{j}+\alpha_{j} \boldsymbol{p}_{j}$
5. $\quad \boldsymbol{r}_{j+1}:=\boldsymbol{r}_{j}-\alpha_{j} \boldsymbol{A} \boldsymbol{p}_{j}$
6. $\beta_{j}:=\left\langle\boldsymbol{r}_{j+1}, \boldsymbol{A} \boldsymbol{r}_{j+1}\right\rangle /\left\langle\boldsymbol{r}_{j}, \boldsymbol{A} \boldsymbol{r}_{j}\right\rangle$
7. $\quad \boldsymbol{p}_{j+1}:=\boldsymbol{r}_{j+1}+\beta_{j} \boldsymbol{p}_{j}$
8. EndDo

This algorithm generates a set of vectors $\left\{\boldsymbol{A} \boldsymbol{p}_{j}\right\}$ which are orthogonal. Since $\boldsymbol{p}_{j}$ belongs to the Krylov subspace $K_{j}$, one can associate canonically with the sequence $\boldsymbol{p}_{j}$ a sequence of polynomials $\pi_{j}$, where $\pi_{j}$ is of degree $j$. The relation between $\boldsymbol{p}_{j}$ and $\pi_{j}$ is $\boldsymbol{p}_{j}=\pi_{j}(\boldsymbol{A}) \boldsymbol{b}$. The sequence of polynomials $\left\{t \pi_{j}(t)\right\}$ is also orthogonal with respect to
the discrete inner product $\langle\boldsymbol{p}, \boldsymbol{q}\rangle_{b}=\left\langle\pi_{p}(\boldsymbol{A}) \boldsymbol{b}, \pi_{q}(\boldsymbol{A}) \boldsymbol{b}\right\rangle$, where we have denoted by $\pi_{p}$ the polynomial associated with the vector $\boldsymbol{p}$ in $K_{m}$.

If we replace the discrete inner product $\left\langle\pi_{p}, \pi_{q}\right\rangle_{b}$ with any inner product $\langle p, q\rangle_{w}$, we will obtain an algorithm which formally constructs a sequence $\left\{t p_{j}\right\}$ of polynomials which are orthogonal with respect to this inner product. The polynomial $t s_{k}(t)$ which is the closest to $\psi$ it the $w$-norm sense can then easily be determined:

$$
\min _{s \in \mathcal{P}_{k}}\|\psi-t s(t)\|_{w}=\left\|\psi-t s_{k}(t)\right\|_{w} \quad \text { with } \quad s_{k}(t)=\sum_{i=0}^{k} \frac{\left\langle\psi, t \pi_{i}\right\rangle_{w}}{\left\langle t \pi_{i}, t \pi_{i}\right\rangle_{w}} \pi_{i}(t)
$$

The filtered Conjugate Residual algorithm proposed in [20] computes the two sequence of polynomials $\left\{\pi_{j}\right\},\left\{\rho_{j}\right\}$ associated with the usual CR algorithm. These would normally be the scalars and polynomials needed for the case when $\psi=1$. They are indicated with a tilde in the following algorithm. In addition, we will need to generate the above quantities which represent the expansions coefficients of the polynomial in the basis $\left\{t \pi_{j}\right\}$. These are denoted by $\alpha_{j}$ in the algorithm.

Thus, the updates to $\boldsymbol{x}_{j}$ use a different coefficient $\alpha_{j}$ than the coefficient $\tilde{\alpha}_{j}$ used to update $\tilde{\boldsymbol{r}}_{j}$. Inputs to the algorithm are a filter function $\psi$, an inner product $\langle p, q\rangle_{w}$, a matrix $\boldsymbol{A}$, and a vector $\boldsymbol{b}$ :

Algorithm 2.3. Filtered Conjugate Residual Polynomials Algorithm
0. Compute $\tilde{\boldsymbol{r}}_{0}:=\boldsymbol{b}-\boldsymbol{A} \boldsymbol{x}_{0}, \boldsymbol{p}_{0}:=\tilde{\boldsymbol{r}}_{0} \quad \pi_{0}=\tilde{\rho}_{0}=1 ; s_{0}=0$
1.

## Compute $\lambda \pi_{0}$

2. For $j=0,1, \ldots$, until convergence Do:

$$
\tilde{\alpha}_{j}:=\left\langle\tilde{\rho}_{j}, \lambda \tilde{\rho}_{j}\right\rangle_{w} /\left\langle\lambda \pi_{j}, \lambda \pi_{j}\right\rangle_{w}
$$

$$
\alpha_{j}:=\left\langle\psi, \lambda \pi_{j}\right\rangle_{w} /\left\langle\lambda \pi_{j}, \lambda \pi_{j}\right\rangle_{w}
$$

$$
\underset{\sim}{\boldsymbol{x}_{j+1}}:=\underset{\sim}{\boldsymbol{x}_{j}}+\alpha_{j} \boldsymbol{p}_{j} \quad s_{\sim+1}=s_{\sim}+\alpha_{\sim} \pi_{j}
$$

$$
\tilde{\boldsymbol{r}}_{j+1}:=\tilde{\boldsymbol{r}}_{j}-\tilde{\alpha}_{j} \boldsymbol{A} \boldsymbol{p}_{j} \quad \tilde{\rho}_{j+1}=\tilde{\rho}_{j}-\tilde{\alpha}_{j} \lambda \pi_{j}
$$

$$
\beta_{j}:=\left\langle\tilde{\rho}_{j+1}, \lambda \tilde{\rho}_{j+1}\right\rangle_{w} /\left\langle\tilde{\rho}_{j}, \lambda \tilde{\rho}_{j}\right\rangle_{w}
$$

9. 

$$
\boldsymbol{p}_{j+1}:=\tilde{\boldsymbol{r}}_{j+1}+\beta_{j} \boldsymbol{p}_{j} \quad \pi_{j+1}:=\tilde{\rho}_{j+1}+\beta_{j} \pi_{j}
$$

Compute $\lambda \pi_{j+1}$
10. EndDo

Here $\pi_{j}(\lambda), \tilde{\rho}_{j}(\lambda)$, and $s_{j+1}(\lambda)$ are polynomials of degree $j$ and the vectors $\boldsymbol{p}_{j}, \tilde{\boldsymbol{r}}_{j}, \boldsymbol{x}_{j}$ are the corresponding sequences of vectors

$$
\begin{aligned}
\boldsymbol{p}_{j} & =\pi_{j}(\boldsymbol{A}) \boldsymbol{r}_{0}, \\
\boldsymbol{r}_{j} & =\tilde{\rho}_{j}(\boldsymbol{A}) \boldsymbol{r}_{0} \\
\boldsymbol{x}_{j+1} & =\boldsymbol{x}_{0}+s_{j+1}(\boldsymbol{A}) \boldsymbol{r}_{0}
\end{aligned}
$$

where $\boldsymbol{r}_{0}=\boldsymbol{b}-\boldsymbol{A} \boldsymbol{x}_{0}$.
The solution vector $\boldsymbol{x}_{j+1}$ computed at the $j$ th step of Algorithm 2.3 is of the form $\boldsymbol{x}_{j+1}=\boldsymbol{x}_{0}+s_{j+1}(\boldsymbol{A}) \boldsymbol{r}_{0}$, where $s_{j}$ is the $j$ th degree polynomial:

$$
\begin{equation*}
s_{j+1}(t)=\alpha_{0} \pi_{0}(t)+\cdots+\alpha_{j} \pi_{j}(t) \tag{2.8}
\end{equation*}
$$

The polynomials $\pi_{j}$ and the auxiliary polynomials $\tilde{\rho}_{j}(t)$ satisfy the orthogonality relations [20],

$$
\begin{equation*}
\left\langle t \pi_{j}(t), t \pi_{i}(t)\right\rangle_{w}=\left\langle t \tilde{\rho}_{j}(t), \tilde{\rho}_{i}(t)\right\rangle_{w}=0 \quad \text { for } \quad i \neq j \tag{2.9}
\end{equation*}
$$

In addition, the filtered residual polynomial $\psi(t)-t s_{j}(t)$ minimizes $\|\psi-t s(t)\|_{w}$ among all polynomials $s$ of degree $\leq j-1$. Note also that if $\phi=1-\psi$, then the polynomial $\zeta_{j}(t)=1-t s_{j}$ minimizes $\left\|\phi-\zeta_{j}\right\|_{w}$ among all polynomials $p_{j}(t) \in \mathcal{P}_{j}$ satisfying $p_{j}(0)=1$.

It is worth remarking that the $\pi_{j}$ generated by Algorithm 2.3 are the orthogonal polynomials associated with the weight $t^{2} w(t)$; this result follows from (2.9). When the FCR algorithm with Laguerre polynomials $L_{n}^{\alpha}(t)$ is used, for example, then the $\pi_{j}$ are the generalized Laguerre polynomials $L_{n}^{\alpha+2}(t)$. This observation will be exploited shortly.

Note that the FCR algorithm has a cost per iteration of three vector saxpy operations and one matrix-vector multiply. All other operations are with polynomials and these are usually negligible. They include two polynomial saxpys (the calculation of $s_{j+1}$ on line 5 is not necessary but has been inserted for clarity), three polynomial inner-products, and one polynomial multiply by $t$. If we write $\psi, \pi_{j}$, and $\tilde{\rho}_{j}$ using as basis the orthogonal polynomials associated with $\langle,\rangle_{w}$, then we can evaluate a polynomial inner-product at a cost of $2 j$; the $t$-polynomial multiply will also incur a cost of $6 j$ (by employing the three-term recurrence relation for the orthogonal polynomials to express $t \pi_{j+1}$ in the orthogonal polynomial basis). In short, the overall cost per loop is $6 n+N z(\boldsymbol{A})+16 j, j$ being the loop index and $N z(\boldsymbol{A})$ the number of non-zeros in $\boldsymbol{A}$. To add a termination test like $\left\|\boldsymbol{x}_{j+1}-\boldsymbol{x}_{j}\right\|<\epsilon$, analogous to line 6 in Algorithm 2.1, would increase the iteration cost by another $2 n$. FCR requires storage of $4 n+3 j$ (the four vectors $\boldsymbol{x}_{j}, \tilde{\boldsymbol{r}}_{j}, \boldsymbol{A} \boldsymbol{p}_{j}$, and $\boldsymbol{p}_{j}$, and coefficients for the three polynomials $\tilde{\rho}_{j}, \pi_{j}, t \pi_{j}$ ). Our bookkeeping suggests that Algorithm 2.3 may be marginally faster but requires marginally more memory than Algorithm 2.1 , provided $j$ remains small compared to $n$.

In order to use FCR to compute $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$, let $\phi=\exp (-\tau t)$. Then, with $\boldsymbol{x}_{0}=\mathbf{0}$, apply the FCR algorithm to get $\boldsymbol{x}_{j}=\sum_{k=0}^{j-1} \alpha_{k} \pi_{k}(\boldsymbol{A}) \boldsymbol{b}$. Finally, compute $\boldsymbol{z}_{j}=\boldsymbol{b}-\boldsymbol{A} \boldsymbol{x}_{j}=$ $\left(\boldsymbol{I}-\boldsymbol{A} s_{j}(\boldsymbol{A})\right) \boldsymbol{b}=\zeta_{j}(\boldsymbol{A}) \boldsymbol{b}$ as an approximation for $\phi(\boldsymbol{A}) \boldsymbol{b}=\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$. If one wants the current estimate for $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ at each step in the iteration, one can replace line (5) in Algorithm 2.3 by

$$
\boldsymbol{z}_{j+1}=\boldsymbol{z}_{j}-\alpha \boldsymbol{A} \boldsymbol{p}_{j}
$$

If $\boldsymbol{A}$ is symmetric with eigen-decomposition $\boldsymbol{A}=\boldsymbol{V} \boldsymbol{\Lambda} \boldsymbol{V}^{T}$, then

$$
\begin{aligned}
\left\|\exp (-\tau \boldsymbol{A}) \boldsymbol{b}-\boldsymbol{z}_{j}\right\|_{2} & =\left\|\phi(\boldsymbol{A}) \boldsymbol{b}-\zeta_{j}(\boldsymbol{A}) \boldsymbol{b}\right\|_{2} \\
& =\| \boldsymbol{V}\left(\phi(\boldsymbol{\Lambda})-\zeta_{j}(\boldsymbol{\Lambda}) \boldsymbol{V}^{T} \boldsymbol{b} \|_{2}\right. \\
& \leq \max _{i}\left|\phi\left(\lambda_{i}\right)-\zeta_{j}\left(\lambda_{i}\right)\right| \cdot\|\boldsymbol{b}\|_{2} .
\end{aligned}
$$

The hope is that if $\left\|\phi(t)-\zeta_{j}(t)\right\|_{w}$ is small over some interval $[a, b]$ containing the spectrum of $\boldsymbol{A}$, then $\max _{i}\left|\phi\left(t_{i}\right)-\zeta_{j}\left(t_{i}\right)\right|$ will also be small, although it is hard to guarantee this since $\|\quad\|_{w}$ is only a least squares norm.

The simplest criterion for stopping the algorithm is to measure the difference between two consecutive iterates, so one can stop whenever $\left\|\boldsymbol{x}_{k+1}-\boldsymbol{x}_{k}\right\|_{2} \leq\left\|\boldsymbol{x}_{k}\right\| \epsilon$, where $\epsilon$ is some tolerance.
2.4. Use of modified orthogonal polynomials. As was already mentioned, an important advantage of the procedure presented in the previous section, relative to the straightforward procedure based on orthogonal expansions presented in Section 2.2, is that it matches exactly the exponential at $t=0$. The FCR procedure is rather general and has been used for $\psi$ functions defined as very general spline functions. One may ask whether or not it is possible to derive an algorithm that is equivalent to Algorithm 2.3, but which resembles Algorithm 2.1. This can be done for the exponential function by exploiting the remark made earlier about the choice of $\alpha$.

In order to minimize $\|\psi-t q(t)\|_{\alpha}$ over all polynomials $q$ of degree $k$, we need a sequence of polynomials of the specific form $t p_{i}(t)$ which are orthogonal with respect to the
inner product $\langle,\rangle_{\alpha}$. At this point we recall an observation made earlier: the sequence of polynomials $\left\{t L_{i}^{\alpha+2}(t)\right\}$ is orthogonal with respect to the inner product $\langle,\rangle_{\alpha}$. This is because for $i \neq j$,

$$
\begin{aligned}
\left\langle t L_{i}^{\alpha+2}, t L_{j}^{\alpha+2}\right\rangle_{\alpha} & =\int_{0}^{\infty} t^{\alpha} e^{-t} t^{2} L_{i}^{\alpha+2}(t) L_{j}^{\alpha+2}(t) d t \\
& =\int_{0}^{\infty} t^{\alpha+2} e^{-t} L_{i}^{\alpha+2}(t) L_{j}^{\alpha+2}(t) d t=0
\end{aligned}
$$

As a result, the least-squares polynomial approximation to $\psi$ is given by

$$
t s_{k}(t)=\sum_{i=0}^{k} c_{i} t L_{i}^{\alpha+2}(t) \quad \text { with } \quad c_{i}=\frac{\left\langle\psi, t L_{i}^{\alpha+2}\right\rangle_{\alpha}}{\left\langle t L_{i}^{\alpha+2}, t L_{i}^{\alpha+2}\right\rangle_{\alpha}}
$$

The following proposition defines a recurrence relation to compute the coefficients $c_{i}$.
Proposition 2.1. Let $\psi(t)=1-e^{-\tau t}$, and let the three-term recurrence of the generalized Laguerre polynomials $L_{i}^{\alpha+2}$ be written in the form (2.6). Define $\sigma_{i}=(i+\alpha+2) / i$ for $i>0, \sigma_{0}=1$. Then the polynomial of degree $k+1$ of the form $t s(t)$ which minimizes the norm $\|\psi-t s(t)\|_{\alpha}$ among all polynomials s of degree $\leq k$ is given by

$$
\begin{equation*}
t s_{k}(t)=\sum_{i=0}^{k} c_{i} t L_{i}^{\alpha+2}(t) \tag{2.10}
\end{equation*}
$$

where $c_{i}$ satisfies the recurrence relation

$$
\begin{align*}
c_{i+1} & =\frac{1}{\sigma_{i+1} \beta_{i+1}}\left[\delta_{i 0}-\frac{\tau}{(1+\tau)^{\alpha+4}}-\alpha_{i} c_{i}-\frac{\gamma_{i}}{\sigma_{i}} c_{i-1}\right], i=0,1, \ldots  \tag{2.11}\\
c_{0} & =\frac{1}{\alpha+2}\left[1-\frac{1}{(1+\tau)^{\alpha+2}}\right] \tag{2.12}
\end{align*}
$$

where $\delta_{i j}$ is the Kronecker symbol and we define $\gamma_{0} c_{-1} / \sigma_{0} \equiv 0$.
Proof. The least-squares polynomial approximation to $\psi$ is given by

$$
\begin{aligned}
t s_{k}(t) & =\sum_{i=0}^{k} c_{i} t L_{i}^{\alpha+2}(t) \quad \text { with } \quad c_{i}=\frac{\tilde{c}_{i}}{d_{i}} \quad \text { and } \\
\tilde{c}_{i} & =\int_{0}^{\infty} t^{\alpha} e^{-t} \psi(t) t L_{i}^{\alpha+2}(t) d t, \quad d_{i}=\int_{0}^{\infty} t^{\alpha} e^{-t}\left(t L_{i}^{\alpha+2}(t)\right)^{2} d t
\end{aligned}
$$

Consider $d_{i}$ first and observe that

$$
d_{i}=\int_{0}^{\infty} t^{\alpha+2} e^{-t}\left[L_{i}^{\alpha+2}(t)\right]^{2} d t=\left\|L_{i}^{\alpha+2}(t)\right\|_{\alpha+2}^{2}
$$

It is known that

$$
\left\|L_{i}^{\alpha}\right\|_{\alpha}^{2}=\frac{\Gamma(i+\alpha+1)}{i!}
$$

So $d_{i}=\left\|L_{i}^{\alpha+2}(t)\right\|_{\alpha+2}^{2}=\frac{\Gamma(i+\alpha+3)}{i!}$.

Next, we take up the task of calculating $\tilde{c}_{i}$ :

$$
\begin{align*}
\tilde{c}_{i} & =\int_{0}^{\infty} t^{\alpha+1} e^{-t}\left[1-e^{-\tau t}\right] L_{i}^{\alpha+2}(t) d t  \tag{2.13}\\
& =\int_{0}^{\infty} t^{\alpha+2} e^{-t} \frac{1-e^{-\tau t}}{t} L_{i}^{\alpha+2}(t) d t \tag{2.14}
\end{align*}
$$

There are two possible ways to proceed from here. One is to attempt to calculate the above integral. This can be done by successive integration by parts using the expression

$$
L_{i}^{\alpha}(t)=e^{t} \frac{t^{-\alpha}}{i!} \frac{d^{i}}{d t^{i}}\left(e^{-t} t^{i+\alpha}\right)
$$

Define the function $\phi_{\tau}(t) \equiv\left(1-e^{-\tau t}\right) / t$. With this, $\tilde{c}_{i}$ becomes

$$
\tilde{c}_{i}=\frac{1}{i!} \int_{0}^{\infty} \phi_{\tau}(t) \frac{d^{i}}{d t^{i}}\left(e^{-t} t^{i+\alpha+2}\right) d t
$$

This integral can be evaluated with a succession of integration by parts leading to results involving successive derivatives of $\phi_{\tau}(t)$. This yields a rather complicated expression for $\tilde{c}_{i}$.

An alternative is to exploit the recurrence relation for the orthogonal polynomials, which we assume is in the form (2.6). It is more convenient to deal with the scaled quantities $c_{i}$ directly rather than the unscaled $\tilde{c}_{i}$. We begin by splitting the calculation into two parts. From (2.13), we have

$$
c_{i}=\frac{1}{d_{i}} \int_{0}^{\infty} t^{\alpha+1} e^{-t} L_{i}^{\alpha+2}(t) d t-\frac{1}{d_{i}} \int_{0}^{\infty} t^{\alpha+1} e^{-(1+\tau) t} L_{i}^{\alpha+2}(t) d t \equiv f_{i}-g_{i}
$$

We seek to establish recurrence relations for $g_{i}$ and $f_{i}$ separately. Consider the case $i=0$. For $i=0$, we have $L_{0}^{\alpha+2}(t)=1$, so

$$
f_{0}=\frac{1}{d_{0}} \int_{0}^{\infty} t^{\alpha+1} e^{-t} d t ; \quad g_{0}=\frac{1}{d_{0}} \int_{0}^{\infty} t^{\alpha+1} e^{-(1+\tau) t} d t
$$

Using the standard definition of the Gamma function, we note that (for $\beta>0$ ),

$$
\int_{0}^{\infty} e^{-t} t^{z} d t=\Gamma(z+1) ; \quad \int_{0}^{\infty} e^{-\beta t} t^{z} d t=\frac{\Gamma(z+1)}{\beta^{z+1}}
$$

Therefore, recalling that $d_{0}=\Gamma(\alpha+3)=(\alpha+2) \Gamma(\alpha+2)$, we have

$$
\begin{equation*}
f_{0}=\frac{1}{d_{0}} \Gamma(\alpha+2)=\frac{1}{\alpha+2} ; \quad g_{0}=\frac{1}{d_{0}} \frac{\Gamma(\alpha+2)}{(1+\tau)^{\alpha+2}}=\frac{1}{(\alpha+2)(1+\tau)^{\alpha+2}} . \tag{2.15}
\end{equation*}
$$

For a general $i$ we have for $f_{i+1}$,

$$
\begin{aligned}
d_{i+1} f_{i+1} & =\int_{0}^{\infty} t^{\alpha+1} e^{-t} L_{i+1}^{\alpha+2}(t) d t \\
& =\frac{1}{\beta_{i+1}} \int_{0}^{\infty} t^{\alpha+1} e^{-t}\left[t L_{i}^{\alpha+2}(t)-\alpha_{i} L_{i}^{\alpha+2}(t)-\gamma_{i} L_{i-1}^{\alpha+2}(t)\right] d t
\end{aligned}
$$

Apart from the scaling by $\beta_{i+1}$, the first term on the right-hand side,

$$
\int_{0}^{\infty} t^{\alpha+1} e^{-t} t L_{i}^{\alpha+2}(t) d t=\int_{0}^{\infty} t^{\alpha+2} e^{-t} L_{i}^{\alpha+2}(t) d t
$$

is either equal to $\Gamma(\alpha+3)=d_{0}$ when $i=0$, or to zero for $i>0$ by the orthogonality of the functions $L_{i}^{\alpha+2}$ and 1. Therefore, we have:

$$
d_{1} f_{1}=\frac{d_{0}-\alpha_{0} d_{0} f_{0}}{\beta_{1}}, \quad d_{i+1} f_{i+1}=\frac{-\alpha_{i} d_{i} f_{i}-\gamma_{i} d_{i-1} f_{i-1}}{\beta_{i+1}} \quad \text { for } i>0
$$

As a notational convenience we define $\sigma_{i}=d_{i} / d_{i-1}$ and note that

$$
\begin{equation*}
\sigma_{i}=\frac{\Gamma(i+\alpha+3)}{i!} \frac{(i-1)!}{\Gamma(i+\alpha+2)}=\frac{i+\alpha+2}{i} . \tag{2.16}
\end{equation*}
$$

With this, the above formulas become,

$$
f_{1}=\frac{1-\alpha_{0} f_{0}}{\sigma_{1} \beta_{1}}, \quad f_{i+1}=-\frac{\alpha_{i} f_{i}+\frac{\gamma_{i}}{\sigma_{i}} f_{i-1}}{\sigma_{i+1} \beta_{i+1}} \quad \text { for } i>0
$$

which we write as a single formula by using the Kronecker symbol and setting $\gamma_{0} f_{-1} \equiv 0$ :

$$
\begin{equation*}
f_{i+1}=\frac{1}{\sigma_{i+1} \beta_{i+1}}\left[\delta_{i, 0}-\alpha_{i} f_{i}-\frac{\gamma_{i}}{\sigma_{i}} f_{i-1}\right] \quad \text { for } i \geq 0 \tag{2.17}
\end{equation*}
$$

We can proceed in a similar way for $g_{i}$ :

$$
\begin{aligned}
d_{i+1} g_{i+1} & =\int_{0}^{\infty} t^{\alpha+1} e^{-(1+\tau) t} L_{i+1}^{\alpha+2}(t) d t \\
& =\frac{1}{\beta_{i+1}} \int_{0}^{\infty} t^{\alpha+1} e^{-(1+\tau) t}\left[t L_{i}^{\alpha+2}(t)-\alpha_{i} L_{i}^{\alpha+2}(t)(t)-\gamma_{i} L_{i-1}^{\alpha+2}(t)\right] d t
\end{aligned}
$$

Now the (undivided) first term in the right-hand side is

$$
\int_{0}^{\infty} t^{\alpha+2} e^{-t} e^{-\tau t} L_{i}^{\alpha+2}(t) d t
$$

Apart from a norm scaling factor, this is the $i$ th expansion coefficient of the function $e^{-\tau t}$ in the Laguerre orthogonal sequence $\left\{L_{i}^{\alpha+2}\right\}$, i.e., with respect to the $i$ th degree polynomial. Specifically, from (2.2), we know that

$$
\frac{\left\langle e^{-\tau t}, L_{i}^{\alpha+2}\right\rangle_{\alpha+2}}{\left\langle L_{i}^{\alpha+2}, L_{i}^{\alpha+2}\right\rangle_{\alpha+2}}=\frac{\tau}{\tau+1}(1+\tau)^{-\alpha-3} \rightarrow\left\langle e^{-\tau t}, L_{i}^{\alpha+2}\right\rangle_{\alpha+2}=\frac{\tau d_{i}}{(1+\tau)^{\alpha+4}}
$$

Therefore,

$$
d_{i+1} g_{i+1}=\frac{1}{\beta_{i+1}}\left[\frac{\tau d_{i}}{(1+\tau)^{\alpha+4}}-\alpha_{i} d_{i} g_{i}-\gamma_{i} d_{i-1} g_{i-1}\right]
$$

Finally, dividing through by $d_{i}$ and rearranging terms,

$$
\begin{equation*}
g_{i+1}=\frac{1}{\sigma_{i+1} \beta_{i+1}}\left[\frac{\tau}{(1+\tau)^{\alpha+4}}-\alpha_{i} g_{i}-\frac{\gamma_{i}}{\sigma_{i}} g_{i-1}\right] \tag{2.18}
\end{equation*}
$$

It is now possible to combine (2.18), (2.17), into a single recurrence formula for $c_{i}$, recalling that that $c_{i}=f_{i}-g_{i}$ :

$$
c_{i+1}=\frac{1}{\sigma_{i+1} \beta_{i+1}}\left[\delta_{i 0}-\frac{\tau}{(1+\tau)^{\alpha+4}}-\alpha_{i} c_{i}-\frac{\gamma_{i}}{\sigma_{i}} c_{i-1}\right], i=0,1, \ldots
$$

where $\delta_{i j}$ is the Kronecker symbol. The recurrence can be started with $i=0$, with the convention that $\gamma_{0} c_{-1} \equiv 0$. The initial value $c_{0}$ is known from (2.15):

$$
c_{0}=\frac{1}{\alpha+2}\left[1-\frac{1}{(1+\tau)^{\alpha+2}}\right]
$$

This completes the proof.

## 3. Convergence and stability.

3.1. Numerical aspects. In practice, the effectiveness of the algorithms hinges on their stability behavior and the rate at which the underlying orthogonal series expansions converge to the exponential function. Recall at this point that the conventional reservation about the Taylor series stems essentially from the fact that on one hand its convergence may require summing a large number of terms, and on the other hand its sum may involve intermediate quantities of large magnitude and/or alternate sign, introducing cancellation errors and stability issues in finite precision arithmetic. Whether the orthogonal series expansions are numerically robust depends on how well they can withstand such drawbacks. In this section, we use a number of figures to explore these issues. We include the Chebyshev series in the discussion for comparison purposes. To keep the figures and the discussion simple, we set $\tau=1$ in (2.2) and (2.4), and denote by $p_{m}(x)$ the partial sum of the series of $e^{-x}$ truncated at length $m$. In the case of the Laguerre series (with $\alpha=0$ ),

$$
p_{m}(x)=\frac{1}{2} \sum_{k=0}^{m} \frac{1}{2^{k}} L_{k}(x), \quad 0<x<\infty
$$

while in the case of the Chebyshev series,

$$
p_{m}(x)=\sum_{k=0}^{m} a_{k} T_{k}\left(\frac{x-l_{2}}{l_{1}}\right), \quad a<x<b
$$

with $l_{1}=(b-a) / 2, l_{2}=(a+b) / 2$, and the coefficients $a_{k}$ as given earlier in (2.5). We first focus on the real case, bearing in mind that we can always scale a nonnegative definite matrix and use a time-stepping procedure to confine the eigenvalues to a fixed interval.


FIG. 3.1. Coefficients and approximation error of the Chebyshev series.
To give the reader a sense of how well the Chebyshev series performs, we consider different choices of the target interval $[a, b]$ and plot in Figure 3.1 the evolution of the coefficients $a_{k}$. It is seen that the coefficients decrease very rapidly in magnitude, and considering that $\max _{x \in[a, b]}\left|T_{k}(x)\right| \leq 1$, we can expect the method to remain stable and converge rapidly for any $x \in[a, b]$. This is indeed what the error curves in the figure show. These observations
corroborate [3, 23]. The error curves also remind us that the target interval has to be chosen carefully to suit the problem at hand, because the approximations degrade rapidly outside their intended interval. Thus, in our matrix exponential context, the eigenvalues need to be strictly inside the interval to prevent spurious effects. Increasing the degree $m$ to take in more terms does not expand the fit interval, but does improve the quality of fit within the target interval. All this agrees with the theory, because the Chebyshev series is, by construction, aimed at a (fixed) given interval. Once constructed, the series does well for any point taken in that interval and it gets better as we use more terms in the series. But the series is not effective outside the target interval for which it is built. Targeting a different interval requires building another series, which is why we need an estimate of the spectral interval for the Chebyshev approximation to be effective.


FIG. 3.2. Laguerre polynomials and approximation error of the Laguerre series.
In the case of the Laguerre series, an immediate contrast with the Chebyshev series is that the Laguerre polynomials $L_{k}(x)$ are not bounded by unity, as can be seen on Figure 3.2, even though the coefficients of the Laguerre series are $1 / 2^{k+1}$. We may wonder if the Laguerre series may be susceptible to numerical difficulties. We see from the figure that this is unlikely because the Laguerre polynomials $L_{k}(x)$ do not oscillate widely with each other in the interval $[0,20]$, which is an interval sufficient for practical purposes (considering the time-stepping strategy hinted to earlier and to be further discussed shortly). Moreover, as the figure shows, their maximum growth stays within $10^{3}$ in this interval, so there is no particular concern in double precision arithmetic. Regarding the speed, we see that the rate of convergence of the Laguerre series is less than that of the Chebyshev series, meaning that if a prescribed accuracy is desired, it takes more terms to achieve that accuracy with the Laguerre series. For example, comparing Figure 3.2 with Figure 3.1, one must go up to $m=40$ terms in the Laguerre series to get a fit comparable to the Chebyshev series with $m=15$ terms for the interval $[0,5]$, or $m=20$ terms for the interval $[0,10]$. The trade-off when computing the matrix exponential is that the Laguerre series does not need an estimate of the bounds of the spectrum as the Chebyshev series does.

As noted earlier, the FCR-based Laguerre series enforces the exactness of the approximation at $x=0$. We turn our attention now to comparing it with the classical Laguerre series.

In Figure 3.3, we can see that both series remain very close around the origin, but as we move away, a difference becomes manifest. The onset of the divergence between the two curves moves to the right as more terms are added. Slight differences can also be seen between the maximum error curves in Figure 3.2 and Figure 3.3. For example in the interval


FIG. 3.3. Variation between the classical Laguerre and FCR-Laguerre, and approximation error of FCR-Laguerre.
$[0,5]$, we can see in the former that the maximum error curve for $m=20$ is slightly under $10^{-6}$, whereas in the latter it is slightly over $10^{-6}$. The discrepancy is a natural consequence of using different optimality criteria for the expansions. However, the differences are too small to be significant in our context, and we shall use either variant in our experiments.


FIG. 3.4. Approximation error of the Chebyshev series in the complex plane.
We now consider the case where the matrix involved may have complex eigenvalues. Although the orthogonal series considered in this study are primarily designed to approximate the exponential of a real variable, it is worth exploring how these series will behave if the real variable $x$ is replaced by a complex variable $z$. This will give us a sense of how the algorithms will behave when used with nonsymmetric matrices that are barely indefinite.

In Figure 3.4 and Figure 3.5, we plot the level curves of the error $\left|e^{-z}-p_{m}(z)\right|$ in


FIG. 3.5. Approximation error of the Laguerre series in the complex plane.
the complex plane using $m \in\{15,20\}$ and $b \in\{5,10\}$ for the Chebyshev series, and $m \in\{15,20,40,50\}$ for the Laguerre series. In the Chebyshev case in Figure 3.4, where $m \in\{15,20\}$, the subplots (a) and (c) give the level curves at a finer resolution, whereas the subplots (b) and (d) aggregate the regions where the error is under $10^{-4}$ and $10^{-5}$, respectively. In the Laguerre case in Figure 3.5, the subplots (a) and (c) use $m \in\{15,20\}$ and aggregate the regions where the error is under $10^{-4}$ and $10^{-5}$, whereas (b) and (d) use the higher degrees $m \in\{40,50\}$ and aggregate the regions under $10^{-12}$ and $10^{-13}$, respectively. These contour plots allow us to contrast the two methods either from the perspective of equal degree or accuracy. We can see that the observations made in the real case apply also in the complex case. For the Chebyshev series, the region of good fit spreads around the target interval. For the Laguerre series, it is anchored at the origin and gradually grows toward the right as more terms are added to the expansion. In both methods, the region only extends slightly in the imaginary direction, or alternatively, it takes more terms to attain a certain accuracy in the complex case. Of note is that an accuracy of $10^{-4}$ or $10^{-5}$ can be achieved by both methods in comparable regions with the same degree $m=15$ or $m=20$, respectively. However, the Chebyshev series is appreciably more accurate around the real axis, as is apparent in the subplots (a) and (c) of Figure 3.4. On the whole, neither method is ideally suited for evaluating the matrix exponential at a high accuracy when the matrix has eigenvalues with large imaginary parts, unless a proper scaling is made to shrink the spectrum. This is discussed next.
3.2. Scaling and staging. This strategy is aimed not only at improving the accuracy by confining the spectrum to a more desirable domain for a given degree of the partial sum, but also at avoiding numerical overflow when summing the terms of the series. It consists of scaling the matrix by a number, say $n_{\text {stage }}$, such that $\|\tau \boldsymbol{A}\| / n_{\text {stage }} \lesssim b$, where $\|\cdot\|$ is some
norm, and $b$ is chosen to target an appropriate interval based on our discussion above. From there, the calculation of $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ is broken into $n_{\text {stage }}$ steps as described below.

## Algorithm 3.1. Scaled and Staged Calculation of $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$

1. Choose $n_{\text {stage }}$ such that $\|\tau \boldsymbol{A}\| / n_{\text {stage }} \lesssim b$
2. $\hat{\tau}=\tau / n_{\text {stage }}$
$\hat{\boldsymbol{A}}=\hat{\tau} \boldsymbol{A}$
$z_{0}=b$
For $t=1: n_{\text {stage }}$
$\boldsymbol{z}_{t}=\exp (-\hat{\boldsymbol{A}}) \boldsymbol{z}_{t-1}$
3. EndFor

This algorithm simply amounts to a (constant) time-stepping strategy as typically used when solving ordinary differential equations. Here, however, we effectively scale the matrix. The implication is that, first, the scaling $\hat{\tau} \boldsymbol{A}=\tau \boldsymbol{A} / n_{\text {stage }}$ shrinks the spectrum (with $\boldsymbol{A}$ assumed symmetric nonnegative definite) into an interval $[0, b]$; and second, subdividing the calculations into these $n_{\text {stage }}$ stages ensures that each stage uses an orthogonal series approximation for $e^{-\hat{\tau} t}$ that converges after relatively few terms in the partial sum.
4. Numerical results. In this section we apply the FCR approach described above to some sample calculations of $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$. All the codes are implemented in MATLAB. We run the series until two consecutive estimates satisfy $\left\|\boldsymbol{z}_{k+1}-\boldsymbol{z}_{k}\right\|_{2} \leq$ TOL, with preset values of the accuracy parameter TOL. We compare with a Krylov method (the expv.m function from the Expokit package [21]), which implements Arnoldi's Full Orthogonalization Method (FOM), and we set 15 as the dimension of the Krylov basis. This is referred to as $\operatorname{Krylov}(15)$ and it does not use the staging-and-scaling described earlier since it has its own built-in time-stepping strategy. Keep in mind, therefore, that the variable $n_{\text {scale }}$ reported in the tables has no bearing in the Krylov method. We supply the same accuracy parameter TOL for the error control criteria there; see [21] for more details. Since the codes are implemented in MATLAB and are not optimized for speed, exact timings are not decisive. Instead, we report the number of matrix-vector products used. (For the purpose of the experiments, we added a counter to that effect in Expokit's expv.m function.) We check the achieved accuracy by reporting $e_{L a g}=\left\|\boldsymbol{z}_{K r y}-\boldsymbol{z}_{L a g}\right\|_{2}$ and $e_{C h e b}=\left\|\boldsymbol{z}_{K r y}-\boldsymbol{z}_{C h e b}\right\|_{2}$, where $\boldsymbol{z}_{K r y}$ is the solution computed by the Krylov method that we use as a reference, while $\boldsymbol{z}_{\text {Cheb }}$ and $\boldsymbol{z}_{\text {Lag }}$ are the approximations obtained by the Chebyshev and Laguerre methods respectively.
4.1. Dielectric waveguide. The matrix used here comes from the Matrix Market collection. It results from a finite difference discretization of the Helmholtz equation that governs a dielectric channel waveguide problem, which arises in many integrated circuit applications. The matrix in the example is of order $n=2048$ with 10,114 non-zero elements. Its sparsity pattern and spectrum are given in Figure 4.1. The spectrum includes complex eigenvalues, but with small imaginary parts $\left(|\operatorname{Im}(\lambda)|<10^{-3}\right)$. The spectrum extends slightly past the origin, suggesting we take $[a, b]=[-1,1]$ in the Chebyshev method. Also, taking $n_{\text {stage }}=1$ is suitable for both the Chebyshev and Laguerre methods. Results for this example are reported in Table 4.1 with $\tau=1, \boldsymbol{b}=(1, \cdots, 1)^{T}$ and TOL ranging from $10^{-2}$ to $10^{-6}$. We see in this example that Expokit's $\operatorname{Krylov}(15)$ detected that one time-step was enough to achieve the desired accuracy. It also appears that the smaller TOL did not induce further computations because the $\operatorname{Krylov}(15)$ solution was already more accurate than requested. For the orthogonal expansion methods, we see that Chebyshev performs quite well. Laguerre achieved the desired accuracy, albeit with more iterations. This agrees with our earlier analysis in Section 3.


Fig. 4.1. Sparsity pattern and spectrum of the Waveguide example.

|  | Matrix-vector products $\left(n_{\text {stage }}=1\right)$ |  | Errors |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: |
| TOL | Krylov(15) | Laguerre | Chebyshev <br> in $[-1,1]$ | $e_{\text {Lag }}$ | $e_{\text {Cheb }}$ |
| $10^{-2}$ | 16 | 16 | 6 | $3.3 \mathrm{e}-03$ | $1.3 \mathrm{e}-03$ |
| $10^{-3}$ | 16 | 19 | 7 | $3.9 \mathrm{e}-04$ | $7.5 \mathrm{e}-05$ |
| $10^{-4}$ | 16 | 22 | 7 | $4.1 \mathrm{e}-05$ | $7.5 \mathrm{e}-05$ |
| $10^{-5}$ | 16 | 26 | 8 | $3.5 \mathrm{e}-06$ | $7.0 \mathrm{e}-06$ |
| $10^{-6}$ | 16 | 30 | 9 | $3.8 \mathrm{e}-07$ | $4.2 \mathrm{e}-07$ |

Results for the Waveguide example. Here $n_{\text {stage }}=1$ everywhere.


FIG. 4.2. Sparsity pattern and spectrum of the Boeing 767 example.
4.2. Boeing 767 matrix. This is an illustrative example in [17]. It arises from modeling a Boeing 767 aircraft with the aim of optimizing its design to suppress flutter of the wings. This involves a nonsmooth, nonconvex optimization method to stabilize a nonsymmetric matrix $\boldsymbol{A}$ that models the aircraft at flutter condition. The resulting stabilized matrix only barely has its spectrum lying in the negative plane. The order is $n=55$ and the eigenvalue closest to the imaginary axis has $\operatorname{Re}(\lambda)=-0.0788$, while the eigenvalue with largest modulus has $|\lambda|=10^{3}$. Figure 4.2 depicts the sparsity pattern and the spectrum, showing that this is a challenging example because it has complex eigenvalues with quite large imaginary parts. Even
more, these eigenvalues are almost purely imaginary, thus making the problem highly oscillatory and difficult for polynomial methods. In addition, the matrix has $\|\boldsymbol{A}\|_{2}=1.69 \cdot 10^{7}$ and $\|\boldsymbol{A}\|_{1}=1.6 \cdot 10^{7}$, so that attempting to infer $n_{\text {scale }}$ using either norm would be very restrictive. The figure shows that a scaling value of $n_{\text {scale }}=160$ is suitable to shrink the spectrum to fit our orthogonal expansion methods.

Since $\boldsymbol{A}$ is stable, we do not need to negate it. We set $\boldsymbol{b}=[1, \cdots, 1]^{T}$ and simply compute $\exp (\tau \boldsymbol{A}) \boldsymbol{b}$ using Algorithm 3.1 with, respectively, $\tau=1,0.1,0.01$ and $n_{\text {scale }}=$ $160,16,1$. We target two values of TOL, $10^{-3}$ and $10^{-6}$. Results are reported in Table 4.2 with $n_{\text {scale }}$ in parentheses. As can be expected, we observe that when $\tau=1$, Chebyshev and Laguerre struggle to cope with this pathological problem. But the table shows better results with reduced values of $\tau$ since they further shrink the spectrum and make the problem more amenable to the polynomial expansion methods.

| TOL $=10^{-3}$ | Matrix-vector products and $n_{\text {stage }}$ |  | Errors |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: |
|  | Krylov(15) | Laguerre | Chebyshev <br> in $[0,10]$ | $e_{\text {Lag }}$ | $e_{\text {Cheb }}$ |
|  |  |  | $5494(160)$ | $2562(160)$ | $9.4 \mathrm{e}-03$ |
| $4.4 \mathrm{e}-02$ |  |  |  |  |  |
| 0.1 | 144 | $533(16)$ | $264(16)$ | $2 . \mathrm{e}-03$ | $7.9 \mathrm{e}-03$ |
| 0.01 | 48 | $34(1)$ | $19(1)$ | $2.7 \mathrm{e}-03$ | $8.4 \mathrm{e}-04$ |
| TOL $=10^{-6}$ |  |  |  |  |  |
| 1 | 1552 | $6880(160)$ | $3218(160)$ | $2.0 \mathrm{e}-04$ | $2.2 \mathrm{e}-04$ |
| 0.1 | 224 | $688(16)$ | $339(16)$ | $1.2 \mathrm{e}-04$ | $1.1 \mathrm{e}-04$ |
| 0.01 | 48 | $43(1)$ | $24(1)$ | $7.1 \mathrm{e}-05$ | $6.3 \mathrm{e}-05$ |

Results for the Boeing example with $\mathrm{TOL}=10^{-3}, 10^{-6}$ and $\tau=1,0.1,0.01$.
4.3. 3D diffusion-convection equations. As a concluding example, we consider the problem of exponential propagation for a discretization of the system

$$
\begin{equation*}
u_{t}=u_{x x}+u_{y y}-\beta u_{x}-\gamma u_{y}, \quad(x, y) \in \Omega \tag{4.1}
\end{equation*}
$$

For our test problem, $\boldsymbol{A}$ is obtained by dividing a square domain $\Omega$ into a uniform $500 \times 500$ mesh and then applying the standard 5-point diffusion-convection discretization

$$
\begin{array}{r}
\frac{d u_{i j}}{d t}=\frac{1}{\delta x^{2}}\left\{4 u_{i, j}-\left(1-\frac{\beta \delta x}{2}\right) u_{i+1, j}-\left(1+\frac{\beta \delta x}{2}\right) u_{i-1, j}\right. \\
\left.-\left(1-\frac{\gamma \delta y}{2}\right) u_{i, j+1}-\left(1+\frac{\gamma \delta y}{2}\right) u_{i, j-1}\right\} . \tag{4.2}
\end{array}
$$

Taking $\beta \delta x / 2=0.2$ and $\gamma \delta y / 2=0.4$ and writing the right hand side of (4.2) as a matrix times a vector $u=u_{i, j}$ yields a $250,000 \times 250,000$ non-symmetric $\boldsymbol{A}$. The $\boldsymbol{b}$ vector is chosen as the initial shape $u_{i, j}(0)=x_{i}\left(1-x_{i}\right) y_{j}\left(1-y_{j}\right)$.

This matrix has $\|\boldsymbol{A}\|_{1}=8$ and its eigenvalue of largest modulus has $|\lambda|<8$, while its spectrum does not extend too far in the imaginary direction $(|\operatorname{Im}(\lambda)|<0.5)$. Hence, taking $n_{\text {stage }}=1$ is suitable, as well as keeping $[a, b]=[0,10]$ in the case of the Chebyshev method as done earlier. Results for this example are reported in Table 4.3, with TOL ranging from $10^{-2}$ to $10^{-6}$. We see in the example that, owing to its automatic step size selection mechanism, Expokit's Krylov(15) uses two steps, which is why its number of matrix-vector products is about twice the size of the Krylov basis (the extra products occurred there as part

|  | Matrix-vector products $\left(n_{\text {stage }}=1\right)$ |  | Errors |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: |
| TOL | Krylov(15) | Laguerre | Chebyshev <br> in $[0,10]$ | $e_{\text {Lag }}$ | $e_{\text {Cheb }}$ |
| $10^{-2}$ | 32 | 20 | 10 | $5.4 \mathrm{e}-05$ | $5.2 \mathrm{e}-03$ |
| $10^{-3}$ | 32 | 24 | 11 | $4.8 \mathrm{e}-06$ | $1.1 \mathrm{e}-03$ |
| $10^{-4}$ | 32 | 28 | 13 | $4.0 \mathrm{e}-07$ | $4.0 \mathrm{e}-05$ |
| $10^{-5}$ | 32 | 31 | 14 | $6.0 \mathrm{e}-08$ | $6.9 \mathrm{e}-06$ |
| $10^{-6}$ | 32 | 35 | 15 | $4.7 \mathrm{e}-09$ | $1.1 \mathrm{e}-06$ |

Results for the diffusion-convection matrix $\boldsymbol{A}$ given by (4.2). Here $n_{\text {stage }}=1$ everywhere .
of the error estimation). As in the first example, it also appears that the smaller TOL did not induce further computations because the two-step solution here was already more accurate than requested. For the orthogonal expansion methods, we see that Chebyshev performs quite well. Laguerre achieved the desired accuracy with more iterations, as expected from our earlier analysis in Section 3. For both of the methods, choosing $n_{\text {stage }}$ is critical to their efficiency. A large $n_{\text {stage }}$ would mean that it takes many more steps to completion than necessary, whereas a smaller $n_{\text {stage }}$ could mean that the actual polynomial approximation to $\boldsymbol{z}_{t}=\exp (-\hat{\tau} \boldsymbol{A}) \boldsymbol{z}_{t-1}$ in Algorithm 3.1 turns out to be of much higher degree than optimal for the job. An optimal choice is that which shrinks the spectrum into a reasonable domain in a way that ultimately leads to an overall small number of matrix-vector products. This is a problem-dependent issue reminiscent of that arising in ODE solvers where one needs to account for stiffness while attempting to use as few steps as possible.
5. Conclusion. Matrix exponential algorithms based on approximating $\exp (-\tau t)$ by a suitable polynomial $p(t)$ have several advantages. Because $\boldsymbol{A}$ only occurs in matrix-vector products, algorithms are matrix-free and can be made independent of the data structure chosen for $\boldsymbol{A}$; the user can 'own' $\boldsymbol{A}$ 's data structure, so to speak, and does not even have to explicitly store $\boldsymbol{A}$ as a matrix. Further, restricting $\boldsymbol{A}$ to matrix-vector multiplies makes it easy to exploit $\boldsymbol{A}$ 's sparsity; no fill-in occurs; and memory usage is capped at $\boldsymbol{A}$ 's storage plus a few vectors.

All these advantages accrue to Algorithm 2.1 and Algorithm 2.3. They accrue also, however, to explicit integration schemes like forward Euler (FE), which solves (1.1) by timestepping with

$$
\begin{equation*}
\boldsymbol{y}(t+h)=\boldsymbol{y}(t)-\int_{t}^{t+h} \boldsymbol{A} \boldsymbol{y}(\tau) d \tau \approx(\boldsymbol{I}-h \boldsymbol{A}) \boldsymbol{y}(t) \tag{5.1}
\end{equation*}
$$

The drawback of FE and other explicit integration schemes, of course, is that the time-step $h$ must be taken very small when $\boldsymbol{A}$ is stiff. In the framework of polynomial methods, FE is based on the approximation

$$
\begin{equation*}
e^{-\tau t} \approx(1-\tau t / n)^{n} \tag{5.2}
\end{equation*}
$$

where $n$ is the number of time steps taken to integrate from $t=0$ to $t=\tau$. The left hand side of (5.2) is stable only if $h=\tau / n$ is taken small enough that $\left|(1-\tau t / n)^{n}\right| \leq 1$.

The question arises whether the more sophisticated polynomial methods we have studied in this paper escape the stability issue faced by explicit integration methods when too large a time-step is taken. Using orthogonal polynomials to compute $\exp (-\tau \boldsymbol{A}) \boldsymbol{b}$ can be thought of as a semi-implicit method for integrating (1.1).

Unfortunately, as we have seen, the FCR method has the property that the number of iterations or stages increases with $\tau \lambda_{\max }$, where $\lambda_{\max }$ is the eigenvalue of largest modulus.

Even when the contribution of $e^{-\tau \lambda_{\max }}$ to the final answer is entirely negligible, the presence of a large $\lambda_{\max }$ in $\boldsymbol{A}$ 's spectrum forces FCR to iterate more. Thus, semi-implicit methods still suffer from the bane of stiffness.

Viewed differently, all polynomial methods require one to approximate $z(t)=\exp (-\tau t)$ by a polynomial over the interval or region that contains the spectrum of $\boldsymbol{A}$. The larger that region is, the more difficult it is (i.e. the higher the degree of polynomial required) to adequately fit $z(t)$ over that region. Staging can be thought of as a way to make that region smaller by replacing each eigenvalue $\lambda_{i}$ by $\lambda_{i} / n_{\text {stage }}$.

With suitable scaling and staging, we have seen that FCR can handle quite stiff systems of ODEs (large spread in eigenvalues) as well as exponentials of unsymmetric matrices. Clearly, scaling and staging are an essential part of the FCR method.

We examined the performance of FCR with the Laguerre and Chebyshev systems of orthogonal polynomials and compared them with the Krylov method. In theory, Chebyshev expansions converge within an ellipse in the complex plane; Laguerre expansions, within a parabola; and Hermite expansions, over the entire plane [5]. Because the individual terms in the series (2.2) and (2.4) can be quite large in magnitude but alternate in sign, one may experience significant loss of precision and even numerical overflow as one tries to compute such expansions with finite-precision arithmetic. However, scaling and staging allow us to overcome these difficulties.

In our tests, we found that Chebyshev usually performed better than Laguerre, though this requires that we be able to localize the spectrum to an interval $[a, b]$. Laguerre does not have this requirement, but one still needs to make a sensible choice of $n_{\text {stage }}$ for both methods.

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    ${ }^{\dagger}$ Mentor Graphics Corporation, Portland, OR 97070 (bernie_sheehan@mentor.com).
    $\ddagger$ Computer Science \& Engineering, University of Minnesota, Twin Cities, MN 55455 (saad@cs. umn. edu).
    ${ }^{\S}$ Department of Mathematics, University of Alabama, Tuscaloosa, AL 35487 (roger.b. sidje@ua.edu).

